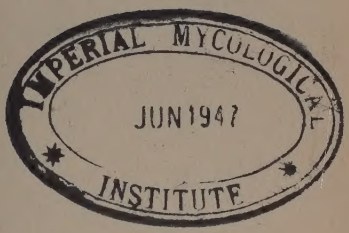
COMPTES RENDUS
(DOKLADY)
DE L'ACADÉMIE DES SCIENCES DE L'URSS

NOUVELLE SÉRIE

1945



VOLUME XLVII

№ 5

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NOUVELLE SÉRIE

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Paraissant tous les dix jours

1945

VOLUME XLVII, № 5

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Édition de l'Académie des Sciences de l'URSS
Moscou 1945

Comptes Rendus (Doklady) de l'Académie des Sciences de l'URSS
1945. Volume XLVII, № 5

MATHEMATICS

CURVES ON CONVEX SURFACES

By A. D. ALEXANDROFF

(Communicated by I. M. Vinogradov, Member of the Academy, 9. X. 1944)

1. The object of the present paper is to outline the fundamental features of a theory that belongs to the inner geometry of surfaces—the theory of curves on convex surfaces. As we subject the convex surfaces under consideration to no restrictions of regularity, the methods of classic analysis so successfully used in the differential geometry fail here, and we have to follow a chain of definitions and reasoning that is entirely strange to the classic method. Within the limits of this paper it is only possible to give the fundamental definitions and results, while those of minor importance and the proofs must be omitted.

By a surface is to be understood here, in the absence of a special statement, any convex surface in a three-dimensional space of constant curvature.

2. Geodesics and the Angle between Geodesics. A line on a surface is said to be the shortest line, if it is shorther than any other of the lines lying on the surface between its two given points. The length of such a line equals the inner distance between its extremities. Every point of a surface has a neighbourhood any two points of which can be joined by the shortest line. By a geodesic is to be understood the line any segment of which, if sufficiently small, is the shortest line.

Let a, b be two geodesics proceeding from a point O, and X, Y the variable points on a and b. Denote by x, y, z the inner distances between O and X, O and Y, X and Y, respectively. Suppose that an angle φ is in existence, defined by the equation

$$\cos\varphi = \lim_{\substack{x\to 0\\ y\to 0}} \frac{x^2+y^2-z^2}{2xy}$$

This angle is taken to be the angle between the geodesics a and b.

The geodesics a and b divide a neighbourhood of the point O into two sectors U and V. If, x and y being arbitrary small, the shortest line XY runs across the sector U, we ascribe to this sector the angle φ. To the sector V we ascribe the angle $\theta-\varphi$, θ being the total angle around the point O, *i. e.* the total angle around the vertex of the cone tangent to the surface at the point O*. If with x and y approaching zero XY runs across U and across V alternately, then $\varphi = \theta-\varphi$, whereby our definition of the angles of sectors contains no contradiction.

Theorem 1. *On a convex surface between any two geodesics a, b, proceeding from a common point O, an angle conformable to the definition made above is in existence.*

* This angle is a quantity belonging to the inner geometry of surfaces.

It is equal to the angle between the tangents to a and b at the point O, as measured on the cone tangent to the surface at O. The angles of the sectors are additive, i. e. if a sector U is the sum of two sectors V and W, the angle of U is the sum of the angles of V and W.*

3. The Direction of a Curve. Let L be a curve on a surface, and O its initial point. Consider an arbitrary geodesic M proceeding from O, and a variable point X on L. Suppose that as X approaches O, the angle between M and the shortest line OX approaches a limit φ. Then we say that the curve L has a definite direction at the initial point O and that this direction forms with M the angle φ (the existence of a direction does not depend on the choice of M). O, if an inner point of L, divides L into two branches; when either branch has a direction at the point O and the angle between these directions equals 180°, the curve L is said to have a direction at O.

Theorem 2. *A curve on a convex surface has a direction at the initial point if and only if it has a tangent at this point.*

If a curve has a direction at an inner point, it has there both a right and a left tangent which do not necessarily coincide; a curve crossing an edge of a polyhedron may serve as an example.

«Direction» is here a substitute for «tangent», but it cannot be substituted with the notion of «tangent geodesic»; thus, the edge of a convex lens has no tangent geodesic, while it has a definite direction at every point. Therefore a convex surface may have such a point from which geodesics cannot be drawn in every direction. Even all points of a convex surface can be as singular as that. But the set of such singular directions around a point always has zero for angular measure.

4. The Integral Geodesic Curvature or the Turning of a Curve. Let L be a geodesic broken line without multiple points. Two sides of it can be distinguished, say, the right and the left side. Let $\varphi_1, \ldots \varphi_n$, be the angles of those sectors between the successive links of L which are lying to the right of L. Define «the right turning» of L by the equation

$$\tau(L) = \sum_{i=1}^{n} (\pi - \varphi_i)$$

The term «right turning» is but an abbreviation of what at full length should read as the curving of the direction of the line L to the right-hand side.

Let now L be a curve without multiple points, having directions at either extremity. We may distinguish two sides of it, say, the right and left. Join the extremities of L by a geodesic broken line P without multiple points lying to the right of L; in other words, the region G bounded by L and P is to lie on the right side of L. Let α and β be the angles between L and P measured in G, and let $\tau(P)$ be the right curving of P. Then the sum $\tau(P) + \alpha + \beta$ is, so to speak, the curving of the direction of L from one extremity to another, but measured along P instead of L. In order to measure it along L, take a sequence of such broken lines P_i convergent to L, and define the right turning of L by the equation

$$\tau(L) = \lim_{i \to \infty} (\tau(P_i) + \alpha + \beta) \qquad (*)$$

The definition of the left curving is quite similar.

Theorem 3. *If a curve without multiple points on a convex surface has definite directions at either extremity, it has both the right and the left curving (i.e. the limit in (*) exists whatever the sequence of P_i). In Euclidian space the sum of these curvings is equal to the area of the spherical image of the curve without the extremities.*

---

* I. Liberman has proved that a geodesic on a convex surface always has a tangent at the initial point [1]. This result of his may be easily extended to convex surfaces in any space of constant curvature.

Consider a simple closed curve L and a point O on L. The curve L may be considered as an open curve L_1 whose extremities coincide at the point O. Suppose the curve L_1 has a definite direction at either extremity, and let φ be the angle between these directions measured, say, on the right side from L_1. Then, $\tau(L_1)$ being the right curving of L_1, the right rotation of the closed curve L is defined by the equation

$$\tau(L) = r(L_1) + (\pi - \varphi)$$

This $r(L)$ does not depend on the choice of the point O, provided it is such that either branch of L has a direction at it.

Theorem 4. *Let a curve L bound on a convex surface in Euclidian space a region G homeomorphic to a circle. Let $\tau(L)$ be the right curving of L supposing the right side to be inward; and let $\omega(G)$ be the area of the spherical image of the (open) region G. Then*

$$\omega(G) = 2\pi - r(L)$$

5. Curves with the Non-negative Rotation. A closed region G is said to be convex in itself, if for any pair of its inner points there is a shortest curve among those lying in G and joining X, Y.

Theorem 5. *A closed region on a convex surface is convex in itself if and only if every segment of its boundary has a non-negative curving to the side of the region.*

Theorem 6. *Let on convex surfaces F_1, F_2,... convergent to F lie the curves L_1, L_2,... without multiple points convergent to a curve L without multiple points. Let the right sides of the curves L_1, L_2,... be defined to converge to the right side of L. If now each of the segments of every curve L_i has a non-negative right rotation, every segment of the curve L has likewise a non-negative right rotation; and the lengths of the curves L_i converge to the lenght of L (every curve each segment of which has non-negative rotation, is rectifiable).*

The right and left rotations of each segment of a geodesic are equal to zero. But, contrary to the state of affairs on regular surfaces, geodesics are here not the only curves having this property. On a convex surface there can exist such curves every segment of which has non-negative and even positive rotation to either side; the edge of a convex lens is an example of such a curve. We call such curves quasi-geodesics. Geodesics are a particular case of quasi-geodesics; and the set of all quasi-geodesics on all convex surfaces is the closure of the set of all the geodesics; the statement is more precisely expressed by the following theorem.

Theorem 7. *If a sequence of quasi-geodesics lying on a convergent sequence of convex surfaces converges to a curve, this curve is a quasi-geodesic on the limit surface. And if L is a sufficiently small segment of a quasi-geodesic on a convex surface F, there exist such a convex surface F' isometric to a neighbourhood of L and such a sequence of convex surfaces F_n convergent to F' that the curve correspondent on F' to L is the limit of a sequence of geodesics lying on the surfaces F_n*.*

This theorem seems to bring to light with sufficient clearness the importance of quasi-geodesics, for we have to expect that many a theorem which in the case of regular surfaces applies to geodesics, must hold for quasi-geodesics when general convex surfaces are taken into consideration. One can, for instance, easily see that even on a closed convex polyhedron the existence of a simple closed geodesic is an exception. But we have strong reason to suppose that the famous Poincaré theorem holds for any closed convex surfaces, if quasi-geodesics are meant instead of geodesics. Further, the well known Morse's theorem asserting the infinity of the number of geodesics joining two points on a closed convex surface may be extended to arbitrary closed convex surfaces, taking quasi-geodesics instead of geodesics.

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Received
9.X.1944.

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* It seems probable that for every quasi-geodesic L on a convex surface F there exist a sequence of convex surfaces F_n convergent to F and a sequence of geodesics L_n lying on F_n and convergent to L.

Comptes Rendus (Doklady) de l'Académie des Sciences de l'URSS
1945. Volume XLVII, № 5

MATHEMATICS

PLANCHEREL'S THEOREM FOR THE GENERALIZED TRANSLATION OPERATOR

By B. LEWITAN

(Communicated by A. N. Kolmogoroff, Member of the Academy, 27.V.1944)

By using D. Raikov's theorem regarding the representation of positive-definite functions on Abelian groups, M. Krein has shown (1) that Plancherel's theorem can be extended to these groups. It will be proved in the present note that Plancherel's theorem is valid for every generalized translation operator (2).

Our proof is different from that of M. Krein.

Theorem. *A unique completely additive measure can be defined on* $\mathfrak{M} - M_0$ *such that the operators*

$$Pf = \int f(t)\,\overline{\varphi(\alpha, t)}\,dm(t), \qquad QE = \int E(\alpha)\,\varphi(\alpha, t)\,dF(\alpha)$$

are defined for all the square-integrable functions $f(t)$, $E(\alpha)$ *(on the corresponding spaces) and are mutually inverse and isometric, i. e.*

$$PQ = QP = 1$$

$$\int |Pf|^2\,dF(\alpha) = \int |f(t)|^2\,dm(t)$$

$$\int |QE|^2\,dm(t) = \int |E(\alpha)|^2\,dF(\alpha)$$

The proof will be carried through step-by-step.

1. Let $f_1(t)$, $f_2(t) \in L_{1,2}$. Let further $x_n(t)$ have the same meaning as in the condition I for generalized translations. Consider the Hilbert space L_2 with the scalar product

$$(x, y) = \int x(t)\,\overline{y(t)}\,dm(t)$$

and put $A_s f = T_t^s f(t)$.

In virtue of the condition III for generalized translations and the formula (2) of our note (3), the relations (3) of the note (4) are fulfilled.

Thus

$$(T_t^s f(t),\ g(t)) = \int T_t^s f(t)\,\overline{g(t)}\,dm(t) = \int \overline{\varphi(\alpha, s)}\,d(E_\alpha f, g) \qquad (1)$$

By the condition V for generalized translations $|\varphi(\alpha, s)| \leqslant c_2$.
Let us show that

$$E_X^{(r)} \tilde{T}_t^r f(t) = E_X^{(t)} \tilde{T}_t^r f(t) = \tilde{T}_t^r E_X f(t) \tag{2}$$

for every set $X \in \mathfrak{M} - M_0$; the upper index at E_X denotes here the variable, which is considered as the principal one (the fact that $\tilde{T}_t^r f(t)$ belongs to L_2 with respect to r follows easily from the conditions IV and V for generalized translations).

In view of the symmetry of T_t^r we have for an arbitrary function $g(t) \in L_2$ keeping in mind that E_X commutes with T^s)

$$E_X^{(r)} T^r g(t) = E_X^{(r)} T_r^t g(r) = T_r^t E_X g(r) = T_t^r E_X g(t)$$

Multiplying the extreme parts of these equalities by $f(t)$ and integrating, we obtain

$$E_X^{(r)} \int g(t)\, \tilde{T}_t^r f(t)\, dm(t) = \int g(t)\, E_X^{(t)} \tilde{T}_t^r f(t)\, dm(t)$$

that proves (2), $g(t)$ being arbitrary.

Put $f(t) = g(t) = x_n(t)$ in the formula (1). We obtain

$$\int T_t^s x_n(t)\, x_n(t)\, dm(t) = \int \overline{\varphi(\alpha, s)}\, d(E_\alpha x_n, x_n)$$

Hence, by formula (2) of the note ([3]):

$$\int \tilde{T}^r T_t^s x_n(t)\, x_n(t)\, dm(t) = \int T_t^s \tilde{T}_t^r x_n(t)\, x_n(t)\, dm(t) =$$
$$= \int \tilde{T}_t^r x_n(t)\, \tilde{T}_t^s x_n(t)\, dm(t) = \int \overline{\varphi(\alpha, r)}\, \varphi(\alpha, s)\, d(E_\alpha x_n, x_n)$$

2. Let X be an arbitrary compactum from $\mathfrak{M} - M_0$. In virtue of the fundamental property of the spectral function E_Δ we have

$$\int \tilde{T}_t^r E_X x_n(t)\, \tilde{T}_t^s x_n(t)\, dm(t) = \int_X \overline{\varphi(\alpha, r)}\, \varphi(\alpha, s)\, d(E_\alpha x_n, x_n) \tag{3}$$

Let us multiply both sides of the equality (3) by $f_1(r)\,\overline{f_2(s)}$ and integrate. We obtain

$$\int dm(t) \int f_1(r)\, \tilde{T}_t^r E_X x_n(t)\, dm(r) \int \overline{f_2(s)}\, \tilde{T}_t^s x_n(t)\, dm(s) =$$
$$= \int_X E_1(\alpha)\, \overline{E_2(\alpha)}\, d(E_\alpha x_n, x_n)$$

where

$$E_i(\alpha) = \int f_i(u)\, \overline{\varphi(\alpha, u)}\, dm(u), \quad (i = 1, 2)$$

Hence by (2), since the ring R is commutative,

$$\int dm(t) \int \tilde{T}_t^r E_X f_1(t)\, x_n(r)\, dm(r) \int \tilde{T}_t^s\, \overline{f(t)}\, x_n(s)\, dm(s) =$$
$$= \int_X E_1(\alpha)\, \overline{E_2(\alpha)}\, d(E_\alpha x_n, x_n) \tag{4}$$

Putting $f_1 = f_2$ and noticing that the left-hand side expression is bounded for all n and $F_n(\Delta) = (E_\Delta x_n, x_n) > 0$, we conclude that the numbers $F_n(X)$ $(n = 1, 2, \ldots)$ are uniformly bounded. Since the set X satisfies the second countability axiom, we may use Helly's theorems regarding the existence of convergent subsequences and proceed to the limit. Let us note that, since the functions $E(\alpha)$ are dense among all the continuous functions on $\mathfrak{M} - M_0$ tending to 0 as $\alpha \to \infty$, the limit of $F_n(\Delta)$ is unique. Denote this limit function by $F(\Delta)$. By passing to the limit as $n \to \infty$ on both sides of the equality (4) we obtain, in virtue of the condition I for generalized translations

$$\int E_X f_1(t)\,\overline{f_2(t)}\,dm(t) = \int_X E_1(\alpha)\,\overline{E_2(\alpha)}\,dF(\alpha)$$

Making $X \to \mathfrak{M} - M_0$, we obtain

$$\int f_1(t)\,\overline{f_2(t)}\,dm(t) = \int E_1(\alpha)\,\overline{E_2(\alpha)}\,dF(\alpha) \tag{5}$$

Since $L_{1,2}$ is everywhere dense in L_2, (5) implies for any $f(t) \in L_2$

$$\int |f(t)|^2\,dm(t) = \int |E(\alpha)|^2\,dF(\alpha)$$

$$E(\alpha) = \text{l. i. m.} \int f(t)\,\overline{\varphi(\alpha, t)}\,dm(t) \tag{6}$$

3. If we substitute in (5) $T_t^s f(t)$ and $x_n(t)$ for $f_1(t)$ and $f_2(t)$, respectively, then

$$\int T_t^s f(t)\,x_n(t)\,dm(t) = \int \varphi(\alpha, s)\,E(\alpha)\,\overline{h_n(\alpha)}\,dF(\alpha)$$

$$E(\alpha) = \int f(t)\,\overline{\varphi(\alpha, t)}\,dm(t), \qquad \overline{h_n(\alpha)} = \int x_n(t)\,\overline{\varphi(\alpha, t)}\,dm(t)$$

As $n \to \infty$, we have

$$f(s) = \int \varphi(\alpha, s)\,E(\alpha)\,dF(\alpha)$$

for every $f(t) \in L$ such that $E(\alpha)$ is also absolutely integrable. It is easily seen that the set of such functions is dense in L_2. For, if $f(s) \in L_2$, then $f(s)$ can be approximated by a function $g(s) \in L_{1,2}$. Now put $g_n(s) = \int T_t^s g(t)\,x_n(t)\,dm(t)$. According to the condition I for generalized translations, g_n tends to g with respect to the norm in L_2. Let us show that the Fourier transform of $g_n(s)$ belongs to L_1.

Indeed

$$\int g_n(s)\,\overline{\varphi(\alpha, s)}\,dm(s) = \int g(s)\,\overline{\varphi(\alpha, s)}\,dm(s) \int x_n(t)\,\overline{\varphi(\alpha, t)}\,dm(t)$$

and, by Schwarz's inequality, is an absolutely integrable function. Thus, putting

$$Pf(t) = E(\alpha) = \int f(t)\,\overline{\varphi(\alpha, t)}\,dm(t)$$

$$QE = \int E(\alpha)\,\varphi(\alpha, t)\,dF(\alpha)$$

we obtain for every function $f(t) \in L_2$

$$QPf = f, \text{ i. e. } QP = 1$$

4. The functions $Pf = E(\alpha) = \int f(t)\,\varphi(\alpha, t)\,dm(t)$, where $f(t)$ are continuous, form an everywhere dense set in L_2. The proof can be carried through as in [1] with the use of formula (6).

This and the equality $PQPf = Pf$ imply $PQG(\alpha) = G(\alpha)$ for every $G(\alpha) \in L_2$, *i. e.* $PQ = 1$. Hence, for every $G(\alpha) \in L_2$

$$\int |G(\alpha)|^2\,dF(\alpha) = \int \overline{G(\alpha)} \left[\int H(t)\,\overline{\varphi(\alpha, t)}\,dm(t) \right] dF(\alpha) =$$

$$= \int |H(t)|^2\,dm(t)$$

$$H(t) = 1.\ \text{i. m.} \int G(\alpha)\,\varphi(\alpha, t)\,dF(\alpha)$$

and the theorem is completely proved.

5. Let Z be a locally compact Abelian group with the second countability axiom, and $m(\Delta)$ be an invariant Haar measure on Z. Put $T_s^* f(t) = f(t+s)$. Let us show in this case that $F(\Delta)$ is an invariant measure on the group of characters. In our case $|\varphi(\alpha, t)| = 1$. Since the topology introduced coincides with the natural topology in the group of characters [2], we have $\varphi(\alpha+\alpha', t) = \varphi(\alpha, t)\,\varphi(\alpha', t)$.

Suppose $f(t)$ is an arbitrary function belonging to $L_{1,2}$. Put

$$E(\alpha) = \int f(t)\,\overline{\varphi(\alpha, t)}\,dm(t), \quad E(\alpha+\alpha') = \int f(t)\,\varphi(\alpha, t)\,\overline{\varphi(\alpha', t)}\,dm(t)$$

By formula (6)

$$\int |E(\alpha+\alpha')|^2\,dF(\alpha) = \int |f(t)|^2\,dm(t) = \int |E(\alpha)|^2\,dF(\alpha)$$

that is

$$\int |E(\alpha+\alpha')|^2\,dF(\alpha) = \int |E(\alpha)|^2\,dF(\alpha) \tag{7}$$

Let $X \in \mathfrak{M} - M_0$ be an arbitrary open set. By choosing $E(\alpha)$ suitably, we can make the right-hand and left-hand sides of (7) to differ arbitrarily little from $F(x)$ and $F(X+\alpha')$, respectively. This proves the invariance of the measure $F(\Delta)$ (the approximating functions are to be chosen in the same way as in [1]).

F. Dzerzhinsky Artillery Academy.

Received 27. V. 1944.

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Comptes Rendus (Doklady) de l'Académie des Sciences de l'URSS
1945. Volume XLVII, № 5

MATHEMATICS

CLASSIFICATION OF SOME SKEW PRODUCTS

By L. PONTRJAGIN, Corresponding Member of the Academy

The present note is devoted to the classification of some fibre-bundles, or skew products, as I prefer to call them, and is based on my papers ([1],[2]). I give here a classification of skew products (see [2], definition 2) $P(R^k, B)$ with the group Γ_k, where R^k is the k-dimensional Euclidian space, Γ_k is the group of all isometric transformations of R^k preserving its origin and orientation, and B is a complex of dimension $n \leqslant 4$. According to the following remark A, the classification depends only on the group Γ_k and the complex B, but not on the space R^k itself. Thus, the results obtained pertain directly to the case considered by Whitney, namely, the skew product $P(S^{k-1}, B)$ with the group Γ_k, where S^{k-1} is the $(k-1)$-dimensional sphere, and Γ_k is its group of rotations. The cases $n=2, 3$ were completely analysed by Whitney ([3]). In the same paper Whitney formulates without proof a result for the case $k=3$, $n=4$.

This result is, however, false, as follows from Theorem 1 and Theorem 2 of the present note.

A. According to propositions A and B of ([2]) the skew product $P(A, B)$ is determined by generating function $\zeta_{\alpha\beta\gamma} \in \Gamma$, $y \in T_\beta$ uniquely, up to isomorphisms. On the other hand, the skew products $P(A, B)$ and $P'(A, B)$ with the same group Γ are isomorphic if and only if the generating functions satisfy the condition: $\zeta_{\alpha\beta\gamma} = \tau^{-1}_{\alpha\gamma}\zeta_{\alpha\beta\gamma}\tau_{\beta\gamma}$, $\tau_{\alpha\gamma} \in \Gamma$. Thus, the classification of skew product depends only on the complex B and the group Γ, but not on the space A.

The universal skew product $P(R^k, H_l)$ (see [2], definition 3 and Theorem 1) will be used for classification of $P(R^k, B)$. Since the dimension n of the complex B is less than or equal to 4, we are interested only in the cycles of dimension $h-r$ in the space $H_l = H(k, l)$, where h is the dimension of H_l and $r \leqslant 4$. Such cycles were found in ([1]). Let us here recall them.

B. Let R^{k+l} be the Euclidian space of dimension $k+l$. Suppose $H(k, l)$ is defined as the set of all the oriented k-dimensional linear spaces in R^{k+l} passing through the origin. Let R^{l+1}, R and R^{l+3} be fixed linear spaces in R^{k+l} of dimension $l+1$, l and $l+3$, respectively, passing through the origin. Denote by z^{h-2} the set of all the R^k in $H(k, l)$, the intersection of which with R^{l+1} is of dimension not less than 2. Denote by z_1^{h-4} the set of all the R^k in $H(k, l)$ whose intersection with R is of dimension not less than 2. Denote by z_2^{h-4} the set of all the R^k in $H(k, l)$ whose intersection with R^{l+3} is of dimension not less than 4. I have proved in ([1]): a) z^{h-2} is a cycle of dimension $h-2$ modulo 2 for $k > 2$, oriented for $k = 2$; b) z_1^{h-4} is an oriented cycle of dimension $h-4$; c) z_2^{h-4} is a cycle of dimension $h-4$ modulo 2 for $k > 4$, oriented for $k = 4$; $z_2^{h-4} = 0$ for $k < 4$.

I have shown in my paper ([4]) that the Lefschetz intersection of the homology class modulo 2 with itself may be defined as the homology class modulo 4. It turns out that in this sense

$$z^{h-2} \times z^{h-2} \sim z_1^{h-4} - 2z_2^{h-4} \pmod 4 \tag{1}$$

C. If f is a mapping of the complex B into H_l, then we denote by u_f^2, u_{1f}^4, u_{2f}^4 the characteristic ∇-cycles (see ([2]), definition 4) that correspond to the cycles z^{h-2}, z_1^{h-4}, z_2^{h-4}, respectively. The relation (1) implies

$$u_f^2 \times u_f^2 \sim u_{1f}^4 - 2u_{2f}^4 \pmod 4 \tag{2}$$

It was established in ([2]) that the homotopical group $\Pi_r(H_l)$ is isomorphic to the homotopical group $\Pi_{r-1}(\Gamma_k)$ (see ([2]), Theorem 2). Hence we can easily construct the homotopical groups $\Pi_r(H_l)$ of dimension $r = 1, 2, 3, 4$.

D. Since Γ_k is connected, $\Pi_1(H_l)$ is trivial. Since $\Gamma_2 \subset \Gamma_k$ forms a one-dimensional basis of homotopics in Γ_k, the two-dimensional sphere $S^2 \subset H_l$ corresponding to Γ_2 forms a two-dimensional basis of homotopics in H_l; thereby S^2 is free for $k=2$ and is of order 2 for $k>2$. As $\Pi_2(\Gamma_k)$ is trivial, $\Pi_3(H_l)$ is trivial as well. $\Gamma_3 \subset \Gamma_k$ is homeomorphic to the three-dimensional projective space; by mapping a three-dimensional sphere onto Γ_3 with degree 2 we obtain a homotopical basis of the space Γ_3. To this generating sphere there corresponds a sphere S_1^4 in $H_l = H(k, l)$ of dimension 4, which forms a homotopical basis in $H(k, l)$ for $k=3$ and is free. S_1^4 is free also for $k>3$, but is no longer a homotopical basis in $H(k, l)$; S_1^4 can only be imbedded in such a basis. There is a sphere S^3 in Γ_4 which, together with the generator in Γ_3 pointed out above, forms a homotopical basis of Γ_4. To the sphere S^3 there corresponds a sphere $S_2^4 \subset H(k, l)$ which, together with S_1^4, form a homotopical basis of $H(k, l)$ for $k \geqslant 4$. The spheres S_1^4 and S_2^4 are independent for $k=4$, while for $k>4$:

$$2S_2^4 - S_1^4 \text{ is homotopic to zero} \tag{3}$$

The spheres S^2, S_1^4 and S_2^4 may be considered as cycles in H_l. It can be calculated immediately that their intersection indices with the cycles pointed out in B have the values

$$I(S^2, z^{h-2}) = 1 \tag{4}$$

$$I(S_1^4, z_1^{h-4}) = 4;\ I(S_1^4, z_2^{h-4}) = 0;\ I(S_2^4, z_1^{h-4}) = 2;\ I(S_2^4, z_2^{h-4}) = 1 \tag{5}$$

By definition 3 and Theorem 1 in ([2]) the problem of classification of the skew products $P(R^k, B)$ is reduced to the classification of continuous mappings of B into $H(k, l)$.

The solution of the latter is given by

Theorem 1. *Let B be a four-dimensional complex and $H(k, l)$ $(k>3,\ l \geqslant 5)$ be the space defined in* B. *Suppose the ∇-cycles u^2, u_1^4, u_2^4 are given in B: u^2 is a two-dimensional ∇-cycle modulo 2; u_1^4 is a four-dimensional ∇-cycle with respect to the field of integral numbers; u_2^4 is a four-dimensional ∇-cycle modulo 2 or integral according as $k>4$ or $k=4$. Suppose that these cycles satisfy the condition*

$$u^2 \times u^2 \sim u_1^4 - 2u_2^4 \pmod 4 \tag{6}$$

There exists then a mapping f of the complex B into $H(k, l)$ such that (see C)

$$u_f^2 \sim u^2;\quad u_{1f}^4 \sim u_1^4;\quad u_{2f}^4 \sim u_2^4 \tag{7}$$

If, moreover, B possesses no four-dimensional even ∇-torsions, and f, g are map-

pings of B into $H(k, l)$ such that

$$u_f^2 \sim u_g^2; \quad u_{1f}^4 \sim u_{1g}^4; \quad u_{2f}^4 \sim u_{2g}^4 \tag{8}$$

then the mappings f and g are equivalent.

When B has four-dimensional even ∇-torsions, then the solution of the problem is not so simple. Formulation of the result and especially the proof become rather complicated, so we would not give them here.

Proof. We shall prove only the second part of the theorem, *i. e.* the equivalence of the mappings f and g under the conditions (8). Denote by B^n the n-dimensional part of the complex B. Let $p \in S^2 \subset H(k, l)$ (see D). The fundamental group of $H(k, l)$ being trivial, we may suppose $f(B^1) = g(B^1) = p$. As the homotopical two-dimensional basis $H(k, l)$ consists only of S^2, we may suppose $f(B^2) \subset S^2$, $g(B^2) \subset S^2$. Let T^2 be an arbitrary oriented two-dimensional simplex in B. Denote by $v_f^2(T^2)$, $v_g^2(T^2)$ the degrees of the mappings f and g of the simplex T^2 onto S^2. Then v_f^2 and v_g^2 are integral two-dimensional ∇-chains in B. The relations (4) and (8) imply $v_f^2 \sim v_g^2 \pmod 2$. By deforming the mapping g in a well-known way, we can make $v_f^2 \equiv v_g^2 \pmod 2$. Hence follows, since S^2 is of order 2, that we can make f and g to coincide on B^2. Since the three-dimensional homotopical group of $H(k, l)$ is trivial, we may suppose f and g coincide already on B^3. Now let us narrowly examine the case $k = 3$. In this case the four-dimensional homotopical basis of $H(k, l)$ consists of S_1^4 only (in case $k > 3$ we have to consider also S_2^4). Let T^4 be an arbitrary oriented four-dimensional simplex in B. Since f and g coincide on the boundary of T^4, they both generate a mapping of an oriented four-dimensional sphere into $H(k, l)$, which is homotopic to a multiple of the sphere S_1^4. Let $v_{fg}^4(T^4)$ denote the multiplicity. Then v_{fg}^4 is a four-dimensional ∇-chain in B. The relation (5) shows that $4v_{fg}^4 = u_{1f}^4 - u_1^4$, while (8) implies $4v_{fg}^4 \sim 0$. This gives, under the condition that B has no four-dimensional even ∇-torsions, $v_{fg}^4 \sim 0$. Hence the equivalence of f and g follows easily. The proof for the case $k > 3$ can be carried through analogously; thereby the sphere S_2^4 presents no difficulty connected with torsions because of the equality $I(S_2^4, z_2^{h-4}) = 1$.

Theorem 1 shows that the characteristic cycles of the skew product $P(R^k, B^4)$ are restricted only by the condition (6). When B^4 is an oriented manifold M^4, and the skew product $P(R^4, M^4)$ consists of tangents, we have moreover

$$u_1^4 \sim u_2^4 \pmod 2 \tag{9}$$

The question whether the restrictions on the characteristic cycles of the manifold M^4 are exhausted by the relations (6) and (9) is rather important for the solution of some geometrical problems. The characteristic cycles u_2^4 and u_1^4 of the manifold M^4 are determined uniquely, up to homologies, by integral numbers; the number corresponding to u_2^4 is the Euler characteristic, while that corresponding to u_1^4 is a new invariant. Calculation of the latter for concrete examples is rather difficult. For the complex projective space of dimension 2 it is equal to 3. It is interesting to note that $u_1^4 \sim 0$ in case M^4 is the boundary of a five-dimensional oriented manifold. I do not know whether the converse is true.

Now we note an elementary result which is proved without using the notion of characteristic cycle.

Theorem 2. *Suppose Γ is a connected topological group, and $P(A, B)$ is a skew product with the group Γ (see [2], definition 1), which is direct on the n-dimensional part B^n of the complex B; this means that there exists a homeomorphic mapping $f(x, y)$ of the direct product $A \cdot B^n$ ($x \in A$, $y \in B^n$) onto the skew product $P(A, B^n)$ such that $f(x, y) = f_y(x) \in A_y$ and $f_{x_0}^{-1} f_{x_1 y} \in \Gamma$, $y \in T_n \subset B^n$ (see [2], definition 1). Let h be a continuous mapping of an oriented n-dimensional sphere S^n into Γ; then $h(S^n)$ is an n-dimensional cycle in Γ. The group of all*

the homology classes that contain cycles of such kind will be denoted by $\sum_n$. *If* T_α^{n+1} *is an arbitrary oriented* $(n+1)$*-dimensional simplex in* B *and* S_α^n *is the boundary of* T_α^{n+1}, *then the mapping* $f_y^{-1}\xi_{\alpha y} \in \Gamma\ (y \in S_\alpha^n)$ *is defined on* S^n, *which generates an element* $w_f^{n+1}(T_\alpha^{n+1})$ *of the group* $\sum_n$. *It turns out that* w_f^{n+1} *is an* $(n+1)$*-dimensional* ∇*-cycle in* B *with respect to the field* $\sum_n$; *the homology class which contains this cycle, is an invariant of the skew product* $P(A.B)$. *i. e. it does not depend on the choice of the mapping* f.

Proof of the theorem is based on the fact that if g is a mapping analogous to f, then $f_y^{-1}g_y = f_y^{-1}\xi_{\alpha y}\ (g_y^{-1}\xi_{\alpha y})^{-1} \in \Gamma$ is a continuous mapping of B^n into Γ.

Received
24. II. 1945.

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Comptes Rendus (Doklady) de l'Académie des Sciences de l'URSS
1945. Volume XLVII, № 5

MATHÉMATIQUES

SUR LES SUITES DES MATRICES STOKHASTIQUES

Par T. A. SARYMSAKOV, de l'Académie des Sciences d'Uzbekistan

La présente Note a pour but d'étendre certains théorèmes de Frobenius (1) et Romanovsky (2) sur matrices non-négatives à une classe spéciale de suites de matrices.

La matrice

$$P = \| p_{ij} \| \quad (i, j = \overline{1, n}; [i, j = 1, 2, 3, \ldots])$$

dont les éléments sont réels, est nommée positive (non-négative), si l'on a pour tous les i et j

$$p_{ij} > 0 \quad (p_{ij} \geqslant 0)$$

ce que l'on écrit dans la forme

$$P > 0 \quad (P \geqslant 0)$$

Suivant Romanovsky, la matrice P est nommée stokhastique, si elle jouit des propriétés:

1° Dans la matrice P aucune ligne et aucune colonne n'est formée de zéros seulement;

2° $\sum_j p_{ij} = 1$ $(i, j = \overline{1, n}; [i, j = 1, 2, 3, \ldots])$;

3° $p_{ij} \geqslant 0$.

La théorie des matrices stockhastiques, sans parler d'autres applications possibles, joue un rôle important dans l'étude des chaînes de Markoff. En particulier, par exemple, l'étude des chaînes de Markoff à une infinité dénombrable d'états par la méthode des matrices conduit à la nécessité d'étudier les propriétés des suites de matrices stokhastiques, ce qui présente un intérêt en vertu de l'importance de cette théorie elle-meme. Le terme général d'une telle suite est de la forme (nous l'écrirons dans la forme qui convient à notre but):

$$\mathfrak{Q}_n = \begin{pmatrix} p_{11} & p_{12} & \cdots & p_{1n} & p_{1\omega_n} \\ p_{21} & p_{22} & \cdots & p_{2n} & p_{2\omega_n} \\ \cdots & \cdots & \cdots & \cdots & \cdots \\ p_{n1} & p_{n2} & \cdots & p_{nn} & p_{n\omega_n} \\ p_{\omega_n 1} & p_{\omega_n 2} & \cdots & p_{\omega_n n} & p_{\omega_n \omega_n} \end{pmatrix}$$

où

$$p_{i\omega_n} = \sum_{j=n+1}^{\infty} p_{ij},\quad p_{\omega_n i} = \frac{\sum_{j=n+1}^{\infty} p_j^0 p_{ji}}{\sum_{j=n+1}^{\infty} p_j^0},\quad p_{\omega_n \omega_n} = \frac{\sum_{j=n+1}^{\infty} p_j^0 p_{j\omega}}{\sum_{j=n+1}^{\infty} p_j^0},\quad (i = \overline{1, n})$$

et toutes les matrices de la suite sont stokhastiques, c'est-à-dire elles vérifient les conditions 1°, 2°, 3°; les nombres $p_j^0 \geqslant 0$ $(j = 1, 2, 3, \ldots)$ sont assujettis aux conditions $\sum_{j=1}^{\infty} p_j^0 = 1$; si $\sum_{j=n+1}^{\infty} p_j^0 = 0$, on pose $p_{\omega_n i} = 0$ pour $i = \overline{1, n}$ et $p_{\omega_n \omega_n} = 1$.

Nous dirons que la suite de matrices stokhastiques est engendrée par la matrice infinie

$$P = \| p_{ij} \| \quad (i, j = 1, 2, 3. \ldots)$$

Introduisons dans nos considérations la suite

$$\Omega_1(\lambda), \Omega_2(\lambda), \ldots, \Omega_n(\lambda), \ldots,$$

où $\Omega_n(\lambda) \equiv | \lambda E_n - \Omega_n |$, et E_n est la matrice unitaire d'ordre n. Pour abréger, nous allons nommer les racines de l'équation $\Omega_n(\lambda) = 0$ racines de la matrice.

Théorème 1. *Toutes les matrices de la suite* $\{\Omega_n\}$ *ont pour racine* $\lambda_0 = 1$; *cette racine est simple et son module surpasse ceux de toutes les autres racines de la matrice* Ω_n *pour chaque* n, *si la matrice* P *d'ordre infini qui engendre la suite* $\{\Omega_n\}$ *est positive.*

En suivant Frobenius, nous dirons qu'une matrice P est décomposable, si elle est de la forme

$$P = \begin{pmatrix} R & Q \\ S & T \end{pmatrix}$$

où R et T sont des submatrices carrées de la matrice P, et Q et S sont, en général, des submatrices rectangulaires et telles que $Q = 0$ ou bien $S = 0$ (le symbole $Q = 0$ signifie que tous les éléments de la matrice Q sont nuls). Si l'on a simultanément $Q = 0$ et $S = 0$, Frobenius dit que la matrice P est «totalement décomposable». Dans le cas où la matrice P est infinie, nous supposerons, d'ailleurs, que la submatrice R est d'ordre fini, tout en restant carrée. Le cas où la submatrice P est d'ordre infini sera considéré dans un autre mémoire.

Les propriétés des matrices décomposables et indécomposables en général et stokhastiques en particulier ont été étudiées d'une manière assez détaillée par Frobenius (1) et Romanovsky (2).

Théorème 2. *Toutes les matrices de la suite* $\{\Omega_n\}$, *sauf, peut être, un nombre fini d'elles, sont simultanément décomposables ou indécomposables.*

Ce théorème permit d'introduire la définition suivante: une suite de matrices stokhastiques $\{\Omega_n\}$ est nommée décomposable, si l'un au moins de ses termes est décomposable.

Théorème 3. *Si la matrice* Ω_n *est de la forme**

$$\Omega_n = \begin{pmatrix} L_{11} & 0 & 0 \ldots 0 \\ 0 & L_{22} & 0 \ldots 0 \\ \cdot & \cdot & \cdot \cdot \cdot \cdot \cdot \\ \cdot & \cdot & \cdot \cdot \cdot \cdot \cdot \\ 0 & 0 & 0 \ldots L_{mm} \end{pmatrix} \qquad (m \leqslant n),$$

les matrices $\Omega_{n+1}, \Omega_{n+2}, \ldots$ *ont la même forme.*

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* L_{kk} $(k = \overline{1, m})$ sont les champs de valeur zéro; tous les autres champs consistent, sans exception, de zéros.

En tenant compte des théorèmes 2 et 3, il est facile d'énoncer, d'une manière convenable les théorèmes III et IV de Frobenius et les théorèmes IX, XII, XIII, XV et XVI de Romanovsky pour le cas de la suite $\{\Omega_n\}$. Par exemple, le théorème XVI de Romanovsky peut être énoncé de la manière suivante.

Théorème 4. *Si pour une certaine valeur de n la matrice Ω_n de la suite est indécomposable, mais pour un entier positif m la matrice Ω_n^m est décomposable, les matrices $\Omega_{n_m}^m$, Ω_{n+1}^m, ... sont totalement décomposables (Ω_n^m est la m-ième puissance de la matrice Ω_n).*

En suivant encore Frobenius et Romanovsky, nous dirons qu'une matrice stokhastique Ω_n est primitive, si toutes ses racines, sauf $\lambda_0 = 1$, ont un module inférieur à 1. Si parmi les racines, excepté l'unité, il y a au moins une dont le module est égal a 1, Ω_n est nommée imprimitive. Une suite de matrices stokhastiques est nommée primitive, si chacun de ses termes est primitif.

Théorème 5. *Si la matrice Ω_1 est imprimitive, toutes les matrices suivantes sont aussi imprimitives.*

Théorème 6. *Si $p_{jj} > 0$ pour une valeur de j au moins, toutes les matrices de la suite $\{\Omega_n\}$ seront primitives.*

Introduisons enfin la définition: La suite $\{\Omega_n\}$ de matrices stokhastique est nommée cyclique d'indice k, si l'une au moins des matrices Ω_n, $n > k$, est cyclique d'indice k *.

Ensuite, il est facile d'étendre les théorèmes XXII—XXVI de Romanovsky [2] au cas de la suite des matrices stockhastiques. Ainsi, par exemple, le théorème XXV de Romanovsky peut etre énoncé dans notre cas de la manière suivante:

Théorème 7. *Si $\Omega_n(0) \neq 0$ à partir d'une certaine valeur de n, aucune des matrices de la suite $\{\Omega_n\}$, à partir du n-ième terme, ne peut être cyclique d'un indice qui soit un diviseur de son ordre. Donc, si $n+1$ est l'ordre de Ω_n et si cet ordre est un nombre premier, Ω_n n'est pas cyclique, ou bien elle est cyclique d'indice $n+1$.*

Dans ce dernier cas chaque matrice Ω_s ($s \geqslant n$) peut etre écrite dans la forme

$$\Omega_s = \begin{Vmatrix} 0 & p_{12} & 0 \dots 0 \dots 0 & 0 \\ 0 & 0 & p_{23} \dots 0 \dots 0 & 0 \\ \dots & \dots & \dots & \dots \\ 0 & 0 & 0 \quad p_{nn+1} & p_{ns} \; p_{n\omega_s} \\ p_{n+11} & 0 & 0 \dots 0 \dots 0 & 0 \\ \dots & \dots & \dots & \dots \\ p_{s1} & 0 & 0 \dots 0 \dots 0 & 0 \\ p_{\omega_{s1}} & 0 & 0 \dots 0 \dots 0 & 0 \end{Vmatrix}.$$

Université d'État
à Tachkent.

Manuscrit reçu
le 1. X. 1944.

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* Pour la définition de cycle voir Romanovsky [2].

Comptes Rendus (Doklady) de l'Académie des Sciences de l'URSS
1945. Volume XLVII, № 5

THEORY OF ELASTICITY

ON THE BOUNDARY ZONE EQUATIONS IN THE THEORY OF SHELLS

By J. N. RABOTNOV

(Communicated by L. S. Leibenson, Member of the Academy, 31. V. 1944)

1. The Simplification of Fundamental Equations. It is well known that in shells concentrated factors and stresses applied at the boundary produce flexure stresses only in the neighbourhood of the boundary or of the point, at which the concentrated factor is applied. The existence of this boundary or local effect makes it possible to simplify the calculation of a symmetrically loaded rotation shell; for which the exact equations may be replaced in the boundary zone by the approximate equations of Geckeler. The remaining part of the shell can be calculated by the momentless theory. The theory of Geckeler's boundary effect may be so generalized as to apply to shells of arbitrary shape under arbtrary boundary conditions, with the only restriction that these boundary conditions ensure the impossibility of the mean surface being flexed.

We start from the equations given in an earlier paper [1]. Let us choose the coordinate system on the surface in such a way that in the boundary zone

$$g_{12}=0,\ g_{11}\approx g_{22}\approx 1$$

The second condition assumes that the domain s of variation of the parameters u, v is small in comparison with the curvature radii. In the sequel we shall verify this assumption. Thus, the order of magnitude of the domain of variation of the quantities characterizing the stress state is s, whereas the order of magnitude of the domain of variation of the quantities determining the metrics is R.

The first two compatibility equations (4.2) coincide with the equations of infinitely small flexures, and the third equation (4.3) differs from the flexure equation by its right-hand side, which cannot be neglected as a consequence of our assumption that the boundary conditions exclude the flexure of the mean surface. We must, therefore, preserve the following compatibility conditions:

$$\begin{aligned}
&\frac{\partial\delta_{11}}{\partial v}-\frac{\partial\delta_{12}}{\partial u}-\left\{\begin{matrix}12\\1\end{matrix}\right\}\delta_{11}+\left(\left\{\begin{matrix}11\\1\end{matrix}\right\}-\left\{\begin{matrix}12\\2\end{matrix}\right\}\right)\delta_{12}+\left\{\begin{matrix}11\\2\end{matrix}\right\}\delta_{22}=0,\\
&\frac{\partial\delta_{22}}{\partial u}-\frac{\partial\delta_{12}}{\partial v}+\left\{\begin{matrix}22\\1\end{matrix}\right\}\delta_{11}+\left(\left\{\begin{matrix}22\\2\end{matrix}\right\}-\left\{\begin{matrix}12\\1\end{matrix}\right\}\right)\delta_{12}-\left\{\begin{matrix}12\\2\end{matrix}\right\}\delta_{22}=0,\\
&b_{11}\delta_{22}+b_{22}\delta_{11}-2b_{12}\delta_{12}=g\left[\frac{\partial}{\partial v}\left(\frac{g}{g_{11}}\gamma_{11,2}\right)-\frac{\partial}{\partial u}\left(\frac{g}{g_{11}}\gamma_{12,2}\right)\right]-\\
&-\frac{g}{g_{11}}\left[\left\{\begin{matrix}11\\2\end{matrix}\right\}\frac{\partial}{\partial v}\left(\vartheta+2\,\frac{\varepsilon_{11}}{g_{11}}\right)-\left\{\begin{matrix}12\\2\end{matrix}\right\}\frac{\partial}{\partial u}\left(\vartheta+2\,\frac{\varepsilon_{11}}{g_{11}}\right)\right]-\\
&-2\left(b_{11}\,b_{22}-b_{12}^2\right)\frac{\varepsilon_{11}}{g}.
\end{aligned}\qquad(1.1)$$

If the order of magnitude of the deformation is $[\varepsilon]$, the order of δ_{ij} will be $[\varepsilon]/h$, if the tension—compression stresses and the flexure stresses are considered to be of one order of magnitude. All terms on the left-hand side of the last equation (1.1) are of order $[\varepsilon]/Rh$, whereas on the right-hand side the last term has the order $[\varepsilon]/R^2$, while the other $[\varepsilon]/s^2$. Hence follows that

$$\frac{[\varepsilon]}{Rh} = \frac{[\varepsilon]}{s^2}, \quad s = \sqrt{Rh} \tag{1.2}$$

For sufficiently small h/R the ratio s/R may be also considered as sufficiently small. Consequently, in differentiating the products of the form εg the derivative of g may be neglected, and we may take $d(\varepsilon g) \approx g\, d\varepsilon$.

In the first two equations (1.1) we may thus omit all terms involving Christoffel's symbols, and in the third equation considerably simplify the right-hand side. We obtain

$$\begin{gathered} \frac{\partial \delta_{11}}{\partial v} - \frac{\partial \delta_{12}}{\partial u} = 0, \quad \frac{\partial \delta_{22}}{\partial u} - \frac{\partial \delta_{12}}{\partial v} = 0 \\ b_{11}\delta_{22} + b_{22}\delta_{11} - 2b_{12}\delta_{12} = -\left(\frac{\partial^2 \varepsilon_{11}}{\partial v^2} + \frac{\partial^2 \varepsilon_{22}}{\partial u^2} - 2\frac{\partial^2 \varepsilon_{12}}{\partial u\, \partial v}\right) \end{gathered} \tag{1.3}$$

The equilibrium equations will be also simplified. Namely, we obtain

$$\begin{gathered} \frac{\partial T_{11}}{\partial v} - \frac{\partial T_{12}}{\partial u} + q_1 = 0, \quad \frac{\partial T_{22}}{\partial u} - \frac{\partial T_{12}}{\partial v} + q_2 = 0 \\ T_{11} b_{22} + T_{22} b_{11} - 2T_{12} b_{12} + \left(\frac{\partial N_1}{\partial v} - \frac{\partial N_2}{\partial u} + q_n\right) = 0 \\ \frac{\partial L_{11}}{\partial v} - \frac{\partial L_{12}}{\partial u} = N_1, \quad \frac{\partial L_{22}}{\partial u} - \frac{\partial L_{12}}{\partial v} = -N_2 \end{gathered} \tag{1.4}$$

2. Reduction to Two Equations. When distributed loads are absent, the equations (1.3) and (1.4) are satisfied, if we put

$$\delta_{11} = \frac{\partial^2 \varphi}{\partial u^2}, \quad \delta_{12} = \frac{\partial^2 \varphi}{\partial u\, \partial v}, \quad T_{11} = \frac{\partial^2 F}{\partial u^2}, \quad T_{12} = \frac{\partial^2 F}{\partial u\, \partial v} \tag{2.1}$$

Using the equations connecting the deformations of the mean surface and the parameters of the curvature variation with stresses and moments, which assume in this case the usual form

$$T_{11} = C(\varepsilon_{22} + \mu\varepsilon_{11}), \ldots, \quad L_{11} = -B(\delta_{22} + \mu\delta_{11}), \ldots, \tag{2.2}$$

we obtain the following system of two equations

$$B\Delta\Delta\varphi = L(F) \tag{2.3}$$

$$\Delta\Delta F = -C(1-\mu^2)L(\varphi) \tag{2.4}$$

The symbol $L(\,)$ means herein the operator $L(f) = b_{11}\frac{\partial^2 f}{\partial v^2} + b_{22}\frac{\partial^2 f}{\partial u^2} - 2b_{12}\frac{\partial^2 f}{\partial u\, \partial v}$. For the lines of curvature $L(f) = \frac{1}{R_1}\frac{\partial^2 f}{\partial v^2} + \frac{1}{R_2}\frac{\partial^2 f}{\partial u^2}$. The shearing forces are

$$N_1 = -B\frac{\partial \Delta\varphi}{\partial v}, \quad N_2 = B\frac{\partial \Delta\varphi}{\partial u} \tag{2.5}$$

3. The Mechanical Interpretation of the Obtained Equations. Let us imagine the shell to consist of two layers, the layer A resisting tension, but offering no resistance to flexure, and the layer B in which elastic moments are developed, but in which there are no stretching or compressing stresses. We shall assume that the flexure of the layer B is described by the same equation as the flexure of a flat plate, namely

$$\Delta\Delta w = q/B \tag{3.1}$$

The distributed load q is the pressure on the shell B by the layer A. The pressure q is connected with the stresses arising in the layer B by the well-known formula

$$q = \frac{T_{22}}{R_1} + \frac{T_{11}}{R_2} \tag{3.2}$$

Combining (3.1) and (3.2), we obtain the equation (2.3). Thus, φ represents the normal displacement of the points of the shell. On the other hand, in the layer A there arises a stress state which in the first approximation may be considered as plane. Beside the elastic deformation, however, which is produced by the exterior forces and is to be determined, in the layer A arise deformations which are caused by the flexure of the layer B. These deformations are $\varepsilon^0_{11} = w/R_1$, $\varepsilon^0_{22} = w/R_2$, $\varepsilon^0_{12} = 0$. The total deformations are then:

$$\varepsilon'_{11} = \varepsilon_{11} + \frac{w}{R_1}, \quad \varepsilon'_{22} = \varepsilon_{11} + \frac{w}{R_2}, \quad \varepsilon'_{12} = \varepsilon_{12}.$$

Substituting them into the compatibility equations, we obtain

$$\frac{\partial^2 \varepsilon_{11}}{\partial v^2} + \frac{\partial^2 \varepsilon_{22}}{\partial u^2} - 2\frac{\partial^2 \varepsilon_{12}}{\partial u\, \partial v} = \frac{1}{R_1}\frac{\partial^2 w}{\partial v^2} + \frac{1}{R_2}\frac{\partial^2 w}{\partial u^2}$$

which is the equation (1.3).

The considerations of the present paragraph cannot be, of course, regarded as the foundation of the method. The mechanical arguments, which are involved in them, must be considered only as an interpretation of results obtained in a different way.

4. The Simplest Examples of Application. For a symmetrical and symmetrically loaded rotation shell the equations (2.3) and (2.4) reduce to

$$\frac{d^4 F}{dx^4} = -\frac{C(1-\mu^2)}{R_1}\frac{d^2\varphi}{dx^2}, \quad B\frac{d^4\varphi}{dx^4} = \frac{1}{R_1}\frac{d^2 F}{dx^2}$$

Introducing instead of φ the quantity $N_2 = B(d^3\varphi/dx^3)$ and eliminating F, we obtain

$$\frac{d^4 N_2}{dx^4} + \frac{3(1-\mu^2)}{R_1^2 h^2} N_2 = 0 \tag{4.1}$$

This is the equation of Geckeler's boundary effect.

For a sphere we have $L(f) = \frac{1}{R}\Delta f$. Eliminating from the system (2.3) and (2.4) φ and putting $\Delta F = p$, we obtain

$$\Delta\Delta p = -\frac{3(1-\mu^2)}{R^2 h^2} p \tag{4.2}$$

This equation falls into two:

$$\Delta p = \pm i\frac{\sqrt{3(1-\mu^2)}}{Rh} p.$$

Putting $\sqrt{3(1-\mu^2)}/Rh = 2k^2$ and transforming the Laplacian to polar coordinates, we find Bessel's equation

$$\frac{d^2 p'}{dr^2} + \frac{1}{r}\frac{dp'}{dr} - \left(\frac{1}{r^2} \pm 2ik^2\right) p' = 0 \tag{4.3}$$

which leads to the well-known solution of the problem on the flexure of a spherical shell by a concentrated force.

Received
31. V. 1944.

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Comptes Rendus (Doklady) de l'Académie des Sciences de l'URSS
1945. Volume XLVII, № 5

PHYSICS

RADIOACTIVE SOURCES OF ELECTRICAL ENERGY

By I. A. LOBANEV and A. P. BELIAKOV

(Communicated by S. I. Wawilow, Member of the Academy, 10. IV. 1945)

A suggestion has recently appeared in the literature (1) to use artificial radioactive substances as a power supply for various electron instruments (electronic microscopes, special types of radio-tubes).

In connexion with this we think that it might be of interest to mention that as early as 1943 we constructed a special electron apparatus in which radioactive substances served to supply the necessary power.

Application of power sources based on the utilization of the electrical energy of radioactive disintegration has yielded positive results. Our nuclear generator of electrical energy ensured constant supply of the apparatus with currents of $10^{-10} - 10^{-9}$ amperes.

The significant successes recorded recently in the production of artificial radioactive elements have opened wide prospects for the use of these elements as sources of electrical energy in various domains of electrical vacuum engineering.

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Received
10. IV. 1945.

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Comptes Rendus (Doklady) de l'Académie des Sciences de l'URSS
1945. Volume XLVII, № 5

PHYSICS

ON NUCLEAR ISOMERS WITH LONG LIFE-TIMES

By L. I. ROUSSINOW and J. M. IGELNITZKI

(Communicated by A. F. Joffé, Member of the Academy, 1.XI.1944)

Atomic nuclei composed of the same elementary particles but with various internal structure—nuclear isomers which were created during the formation of the elements of the periodic system, may still be in existence at this date if their life-time in the metastable state is of the order of $\sim 10^9$ or more years. Such life-times are possible if the excitation energy of the metastable state is small, or if the difference between the mechanical moments of isomeric nuclei in the excited and ground states is large ([1], [2]).

At present a number of formulae which permit one to estimate the lifetime of nuclear isomers with respect to radiation transitions have been derived on various assumptions concerning the model of the radiating system, *i. e.* the nucleus. The life-time computed according to these formulae should then be decreased, due to the possibility of discharge of the isomers by internal electron conversion which must be additionally calculated by special and somewhat complicated formulae.

In the present case one may use L. D. Landau's* formula to estimate the life-time of nuclear isomers; this formula was derived by assuming that the nucleus radiates as a «one-particle» model, and it differs from the other formulae in that allowance is made for the influence of internal electron conversion on the life-time of the isomers.

Landau's formula has the form

$$\tau_{el} = 10^{-12}\,[1 \cdot 3 \cdots (2\Delta l + 1)]^2 \left(\frac{1.5 \cdot 10^{-9}}{\Delta E}\right)^{\Delta l - 1}$$

$$\tau_{magn} = \frac{10^8}{\Delta E}\,\tau_{el}$$

where τ_{el} and τ_{magn} are the life-times of the nuclear isomers in seconds for radiation of electric or magnetic type; ΔE is the excitation energy of the nuclei in eV; Δl is the change of the angular momentum of the nucleus in $\hbar$ units. The value $5 \cdot 10^{-13}$ cm has been accepted for the effective radius of the nucleus.

For the effective radius of the nucleus we put the value $5 \cdot 10^{-13}$cm.

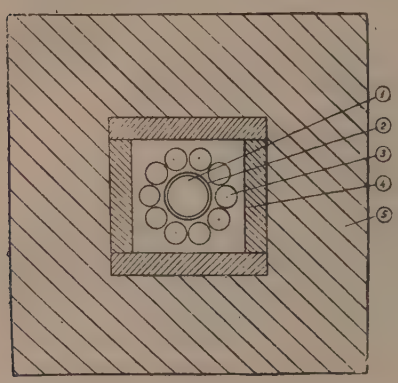
According to Landau's formula, in the case of radiation of the magnetic type for an excitation energy of $2 \cdot 10^3$ eV and a difference of the angular momenta of $4\hbar$, the life-time of the nuclear isomers may be of the order of $\sim 10^9$ years; for an excitation energy of the nuclei of $2 \cdot 10^4$ eV and a difference of the angular momenta of $5\ \hbar$ the life-time of the isomers is $\sim 10^8$ years for electric radiation, and $\sim 10^{12}$ years for magnetic radiation; if the corresponding values are $3 \cdot 10^5$ eV and $6\ \hbar$,

* We wish to express our thanks to Prof. L. D. Landau for permitting us to cite this unpublished formula.

one gets $\tau_{el} \cong 10^9$ years and $\tau_{magn} \cong 10^{12}$ years. At lower values of the excitation energy of the metastable state the life-time of the isomers correspondingly increases.

In spite of the qualitative character of these computations, they show that in a number of cases there exist such values of ΔE and Δl as are sufficient for the existence of nuclear isomers of large life-times.

The aim of the present investigation is to detect experimentally the phenomenon of long-life isomery by the weak electromagnetic rediation (X-rays emitted during internal conversion or γ-rays emitted directly by the nuclei) which accompanies the transition of the excited isomer to the ground state (other methods of detecting this phenomenon, for instance, by the difference in the values of their spins or atomic weights, are surrounded with great difficulties).



1, Geiger-Müller counter filled with a mixture of krypton and xenon; *2*, investigated element; *3*, Geiger-Müller counters coupled to an anticoincidence circuit; *4*, iron filter; *5*, lead block

For the purpose mentioned above a special method was employed. The gamma-quanta were measured by means of an aluminium Geiger-Müller counter 45 mm in diameter and 86 mm long. In order to improve the efficiency of counting of soft gamma-rays the counter was filled with a mixture consisting of 95 per cent of krypton and 5 per cent xenon at a pressure of about 2 atmospheres. The thickness of the counter walls was approximately 0.1 mm. Greater mechanical strength was attained by incorporating three rings, 3 mm wide and 1 mm thick, in the central part of the counter cylinder at equal distances from the edges. The voltage on the counter was near to 5000 volts. The constant counting region was approximately 100 volts broad. This Geiger-Müller counter was surrounded by cylinders of the investigated elements prepared from chemically pure specimens.

In order to increase the sensitivity of the apparatus special measures were taken to decrease the number of counts caused by the background in the counter itself. The central Geiger-Müller counter filled with krypton was surrounded by 10 brass counters, 20 mm in diameter, 100 mm long and with 1 mm walls. These counters were filled with air and a certain amount of argon to a pressure of about 100 mm Hg. The voltage applied to these counters was approximately 2000 volts. The central wires of each of the surrounding counters were connected together, earthed through a common resistance and coupled to an anticoincidence circuit of the amplifier; the main counter was coupled to a second channel of the amplifier. The whole counter system was surrounded by an 20 mm iron absorber and by a 120 mm lead block (see figure). The latter screened the central counter from radiations due to radioactive contaminations in the laboratory and the soft component of cosmic rays.

The iron layer absorbed part of the rays emitted by radioactive contaminations in the lead. Hard rays of the cosmic radiation which actuated the main counter and any one of those surrounding it were not registered because of the action of the anticoincidence circuit. Only pulses in the central counter itself were recorded. Owing to these improvements about 9 discharges per minute in the Geiger-Müller counter were recorded, the background of the counter thus being decreased by over 10 times

Further, by employing Johnson's formulae and considering that absorption of the electromagnetic radiation takes place in the counter gas, we plotted the sensitivity of the counter to gamma-quanta for various wave-lengths. This sensitivity is different for different wave-lengths and can be represented by a curve with minima in the neighbourhood of the absorption edges of Kr and Xe. For 20 keV

gamma-quanta the counting yield is $\sim$100 per cent; for 30 keV, $\sim$40 per cent; for 40 keV, $\sim$25 per cent; for 50 keV, $\sim$12 per cent.

Measurements with the Geiger-Müller counter were carried out in a series of experiments: the investigated element and background of the counter were alternately measured over intervals of 10—20 minutes. Over 10 000 counts were recorded for each of the investigated elements.

Special attention was directed to the investigation of chemically pure elements free from radioactive contaminations. In each case in which additional activity was detected, care was taken to discover the nature of this activity. For this purpose measurements were carried out on the same apparatus with a Geiger-Müller counter filled to $\sim$100 mm Hg with air and quite insensitive to gamma-rays. Furthermore, in order to find out if the observed activity was due to fast electrons or soft gamma-rays, the radiation was absorbed by screens of various atomic numbers.

In the first series of measurements reliable results were obtained for four elements: Cd, In, Au, and Bi; within the accuracy of the statistical errors no soft gamma-rays resulting from long-lived nuclear isomers were observed.

When determining the limits of the life-times of the Cd and In isomers, it was assumed that their excitation energy in the metastable state is higher than the binding energy of the K-electrons of these atoms (26.7 and 27.9 keV), in which internal conversion accompanied by the characteristic K-radiation may take place. In the calculations of τ the sensitivity of the Geiger-Müller counter to this characteristic rediation was taken into account. For Au and Bi it was assumed that the excitation energy of the nuclei was higher than the excitation potential of the L-levels (14.4 and 16.4 keV). The sensitivity of the counter to the characteristic radiation of the L-levels of these elements (and also the Auger effect) were taken into account in the calculations of τ.

If, however, the excitation energy of the nuclear isomers was found to be below the excitation potential of the atoms cited above, and internal conversion in these shells was impossible, the isomer nuclei would radiate gamma-rays directly on transition from the metastable to the ground state.

The order of magnitude of τ in this case would remain the same.

The results for the life-times τ of the investigated elements are given in the table.

Element	τ
Cd_{48}	$> 4 \cdot 10^{15}$ years
$In_{49}^{113,115}$	$> 3 \cdot 10^{15}$ »
An_{79}^{197}	$> 7 \cdot 10^{14}$ »
Bi_{83}^{209}	$> 1 \cdot 10^{15}$ »

(for Cd τ is given for all of its isotopes).

With the method developed here measurements should be carried out with other elements with the aim of discovering long-life nuclear isomers.

In conclusion we wish to express our sincere appreciation to I. V. Kurchatov for interest in this investigation.

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Received
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Comptes Rendus (Doklady) de l'Académie des Sciences de l'URSS
1945. Volume XLVII, № 5

OCEANOLOGY

MAXIMUM MOLECULAR WATER CAPACITY OF BOTTOM DEPOSITS OF THE CASPIAN SEA

By S. W. BRUJEWICZ

(Communicated by W. I. Vernadsky, Member of the Academy, 21. VII. 1944)

As the chemical and physical properties of marine sediments are closely related to their mechanical composition, the determination of the latter in the course of studying the bottom deposits is quite imperative. In order to obtain comparable data it is sufficient to know the mechanical composition from analyses of preliminarily dried material, which is usually done in oceanographic practice. When, however, an exact picture is desired of the distribution of sediments within individual regions of the sea and along the vertical within the sediments, it is necessary to determine the mechanical composition of natural, undried material, since a preliminary drying of the bottom deposits involves a considerable increase in the size of the particles of the marine sediments (adhesion). The content of the «fine fraction» (particles smaller than 0.01 mm), is, however, undervaluated by analysis, and the ground appears to have a coarser composition than it actually has. Thus, from the data of Correns ([3]), in the course of a mechanical analysis of a sample of red clay from the south Atlantic, having a natural water content, there were found particles with an equivalent radius: $>\mu$, 24.4 per cent; $<1\mu$, 75,6 per cent. The same in a sample of marine sediments dried at 105° C: $>1\mu$, 40.2 per cent; $<1\mu$, 59.8 per cent.

As a result of this, Correns recommends when studying recent sediments to carry out the mechanical analysis in samples with a natural water content. Unfortunately, all the methods for determining the mechanical composition of natural or dried samples of sediments are rather cumbersome. On the other hand, one and the same fraction may contain both finer and coarser particles, corresponding to different magnitudes of the inner surface of the ground. Therefore, a fractional mechanical analysis does not give a sufficiently accurate idea of the inner surface of the ground, this magnitude being the index of the degree of its dispersion. These considerations when applied to subaerial soils have long since caused the tendency to substitute for a direct determination of the mechanical composition another, more rapid determination of some or other property of the soil or ground, which is in a close functional relationship with the mechanical composition or, to say more precisely, with the magnitude of the inner surface of the ground. Thus, Mitscherlich ([4]) in 1901 proposed to replace the mechanical analysis by a determination of the maximum hygroscopicity of the soils. In America, Briggs and McLane ([5]), to characterize the mehanical composition of soils, introduced the determination of the «water equivalent», *i. e.* the water content of a soil subjected for 40 minutes to a centrifuge treatment, 1000 *g* (*g*, acceleration of gravity) being developed (the water content was determined at 105—110° C).

With this object in view, A. F. Lebedev proposed the determination of the «maximum molecular water capacity» (m. m. w. c.)—the amount of water corresponding to the maximum moistening of a soil and depending on the capacity of the soil to retain water owing to the molecular forces of cohesion between the soil and water particles. The principle of A. F. Lebedev, that is, a determination of m. m. w. c. by the method of film equilibrium ([2]), was applied in characterizing the mechanical composition of the bottom deposits of the Caspian Sea investigated by the present author. The procedure used is essentially as follows.

The sample of marine sediments, if the natural water content of the ground is found to be insufficient, is moistened with water until a mud condition is reached. Then, upon sheets of filter paper a metallic plate is placed, in the middle the plate has a round hole, measuring 4—5 cm in diameter. Two such plates should be provided, one 2 mm, the other 1 mm thick. The bottom deposits, smoothened if necessary, are transferred to the hole in the plate, being carefully spread all over the hole, after which the plate is removed, and upon the paper a round piece of the sample is left, which is covered from above by a sheet of filter paper. Five or six such samples, separated from one another on each side by 20 sheets of filter paper and oak (10 mm) or metallic spacers, are placed in a hydraulic or simple press (a pressure of about 60 kg per cm^2 is sufficient). For muds and sandy muds the thickness of the layer of bottom deposits is 2 mm; and the time of pressing, 2 minutes; for coarser sediments, the thickness of the layer is 1 mm, and the time, 10 minutes.

After the pressing has been completed, the sample is cleaned from the adhering paper fibres and the water content at 105—110° is determined. The value of the water content referred to dry matter expressed in per cent represents the maximum molecular water capacity.

The values thus determined very closely approach the values of the water content of sediments centrifuged during 2 minutes in a centrifuge at rate of 50 000 revolutions per minute, the acceleration being 70 000 *g*.

Lebedev gives the following average values of the m. m. w. c. for different marine sediments: coarse sand (1—0.5 mm), 1.57 per cent; medium sand (0.5—0.25 mm), 1.60 per cent; fine sand (0.25—0.1 mm), 2.73 per cent; very fine sand (0.1—0.05 mm), 4.75 per cent; silt (0.05—0.005 m), 10.18 per cent; clay (<0.005—0 mm), 44.85 per cent.

The author's determinations of the m. m. w. c., carried out jointly with A. I. Belev ([1]), for the upper layer of the bottom deposits of the Tyub-Karagan Bay, northern part of the Caspian Sea, gave the following values:

	Mud	Sandy mud	Muddy sand	Sand
Maximum molecular water capacity (m. m. w. c.), per cent to natural sediments	12.9	11.3	3.8	1.3
Same, per cent to dry sediments	44.3	22.4	5.0	1.5

It is of interest to note that the m. m. w. c. in per cent to natural sediments for sandy muds and muds with a highly variable mechanical composition is an almost constant value: for the sediments of the Tyub-Karagan Bay, about 13.5; for the upper layers, about 12.5; on the average, about 13 per cent. The water content of the bottom sediments varied from 36 to 77 per cent, and the m. m. w. c. in per cent from dry sediments, from 18 to 48 per cent.

The relations between the m. m. w. c. in natural sediments, the m. m. w. c. in

dry sediments and the water content of natural sediments (H_2O per cent) are expressed by the following formulae:

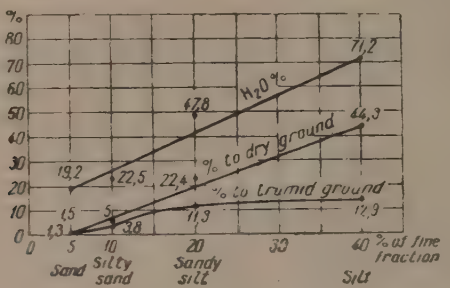
$$\text{m. m. w. c.}_{\text{nat.}} = \text{m. m. w. c.}_{\text{dry}} \left(1 - \frac{H_2O\%}{100}\right) \cong \text{const} \approx 13 \qquad (1)$$

$$\text{m. m. w. c.}_{\text{dry}} \frac{\text{m. m. w. c.}_{\text{nat.}}}{1 - \frac{H_2O\%}{100}} = \frac{1300}{100 - H_2O\%} \qquad (2)$$

It should be noted that the multiplier $\left(1 - \frac{H_2O\%}{100}\right)$ represents the content of dry matter in a unit of weight of natural ground.

Conclusions

1) The determination of the maximum molecular water capacity of natural marine sediments is a reliable index of the magnitude of the inner surface of the ground; it has the advantage of a considerably greater simplicity and speed as compared to the determination of the mechanical composition.



Relation of moisture content ($H_2O\%$) and maximum molecular water capacity (m. m. w. c.) to mechanical composition of sediment (as illustrated by samples of sediments taken with a bottom-sampler from the bottom of the Tyub-Karagan Bay, northern part of the Caspian Sea)

At the same time the determination of the m. m. w. c. does not in all cases replace the determination of the mechanical composition, since the latter gives in a certain, rather rough, degree the spectrum of distribution of the particles of grounds according to their size.

2) Within a rather wide range of variation of the water content and mechanical composition—for muds and sandy muds —the m. m. w. c. in per cent to natural ground is a nearly constant value (about 13), equal to the product of the m. m. w. c. in per cent of dry ground multiplied by the content of dry matter in the marine sediments (formula (1)).

3) Thus, quantitatively the m. m. w. c. $_{\text{dry}}$ is universely proportional to the dry matter content in bottom deposits.

4) Graphically, the water content of natural ground and the m. m. w. c. in per cent to dry sediments as a function of the mechanical composition (content of the fine fraction) is represented almost by straight lines. This is in good agreement with the statement of A. F. Lebedev experimentally proved by him: *viz* «the m. m. w. c. of a soil is an additive function of the mechanical elements of the given soil».

And it is the fine fraction that is the carrier of the water-retaining capacity of sediments.

5) The constancy of the product of multiplication of the m. m. w. c. $_{dry}$ by the content of dry matter in a bottom deposit permits of an approximate estimation of the water content of a marine sediment from the value of the m. m. w. c. $_{dry}$, and reversely.

6) Similarly, from a diagram preliminary constructed for the given grounds it is possible to determine approximately the per cent of the fine fraction from the value of the m. m. w. c. $_{dry}$.

Received
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Comptes Rendus (Doklady) de l'Académie des Sciences de l'URSS
1945. Volume XLVII, № 5

PHYSIQUE APPLIQUÉE

LE PROBLÈME DE WISCHNEGRADSKI DANS LA THÉORIE DE LA RÉGULATION DIRECTE

Par A. ANDRONOW et A. MAYER

(Présenté par L. I. Mandelstam, de l'Académie, le 22. XI. 1944)

Le problème fondamental de la théorie de la régulation directe a été posé par Wischnegradski en 1876 [1]*, en tenant compte de la viscosité et du frottement coulombien. Nous l'appellerons problème de Wischnegradski. Mais Wischnegradski ne le résolut que pour le cas linéaire où le frottement coulombien est absent. Un autre cas particulier du problème de Wischnegradski, tenant compte du frottement coulombien et négligeant le frottement visqueux, a été étudié par Lecornu, Mises, Joukowski. Nous l'appelons problème de Mises. Quoique la théorie d'un grand nombre de dispositifs de régulation, dont certains tout à fait modernes, se ramène après idéalisation convenable au problème de Wischnegradski **, celui-ci, à ce que nous sachons, n'a pas encore été résolu dans sa forme générale.

Nous avons pu récemment donner la solution du problème de Mises en le ramenant à une transformation ponctuelle, d'un plan en lui-même [6]. La même méthode permet de résoudre le problème de Wischnegradski. La marche générale des raisonnements et les calculs eux-mêmes diffèrent peu de ceux, déjà publiés, qui ont trait au problème de Mises.

C'est pourquoi nous nous contenterons dans cette note d'exposer les résultats de notre analyse du problème de Wischnegradski.

1. Les équations simultanées du mouvement du système machine — régulateur peuvent être écrites après les simplifications usuelles (y compris l'hypothèse que sans régulateur la machine n'est pas un système auto-réglable), compte tenu des frottements visqueux et coulombien (ce dernier rend le système non-linéaire), sous la forme:

$$\frac{T^2}{4}\frac{d^2\xi}{d\tau^2} = -\delta\xi + \eta - \frac{T_1}{2}\frac{d\xi}{d\tau} \pm \frac{\varepsilon}{2}; \quad T_0\frac{d\eta}{d\tau} = -\xi \tag{1}$$

où τ est le temps; ξ caractérise le déplacement du manchon; η — la perturbation de la vitesse angulaire; T — la durée de chute libre du manchon; T_0 — la durée de mise en marche de la machine; T_1 — la durée de chute du manchon

* Ce mémoire a joué, on le sait (v., par exemple, [2–4]) un rôle exceptionnel dans l'histoire de la théorie de la régulation: l'inégalité célèbre $AB > 1$ (v. plus bas), cas particulier des conditions générales de Routh-Hurwisz, établies plus tard, a passé de ce travail dans les calculs des ingénieurs.

** V., par exemple, l'ouvrage de Schmidt [5].

en cas de mouvement uniforme dont la vitesse est déterminée par le frottement visqueux; δ — le coefficient d'irrégularité; ε — le coefficient d'insensibilité.

Complétant, compte tenu des propriétés du frottement coulombien, les équations (1) pour le cas où $d\xi/d\tau = 0$ et faisant le changement de variables

$$\tau = \frac{1}{2}\sqrt[3]{T^2 T_0}\, t;\ \xi = \varepsilon \sqrt[3]{\frac{4T_\theta^2}{T^2}}\, x;\ \eta = \varepsilon y \tag{2}$$

nous obtenons le système*:

$$\frac{dx}{dt} = z,\ \frac{dy}{dt} = -x;\ \frac{dz}{dt} = F(x, y, z) \tag{3}$$

où

$$F(x,y,z) = \begin{cases} -Ax + y - Bz - \frac{1}{2} & \text{si } z > 0 \text{ ou } z = 0,\ -Ax + y > \frac{1}{2} \\ -Ax + y - Bz + \frac{1}{2} & \text{si } z < 0 \text{ ou } z = 0,\ -Ax + y < -\frac{1}{2} \\ 0 & \text{si } z = 0,\ -\frac{1}{2} \leqslant -Ax + y \leqslant +\frac{1}{2} \end{cases}$$

Le système (3) et la condition de continuité des fonctions $x(t)$, $y(t)$, $z(t)$ aux points de discontinuité de $F(x, y, z)$ déterminent le problème dynamique à étudier. La théorie classique de la régulation s'intéresse principalement à ce qui se passe dans le système machine — régulateur après une variation brusque de la charge. Les coordonnées x, y étant mesurées à partir du nouveau régime stationnaire, la variation de charge est caractérisée par la position initiale du point figuratif dans l'extension en phase. Dans le cas d'un régulateur centrifuge ordinaire ces conditions initiales peuvent s'écrire: $x = x_0$, $y = Ax_0$, $z = \frac{dx}{dt} = 0$ **.

2. En plus de l'extension en phase nous considérerons encore un plan à coordonnées cartésiennes A, B***. La théorie des transformations ponctuelles des surfaces permet, A et B étant fixés, d'établir la structure qualitative de la division de l'extension en phase x, y, z en trajectoires et les relations quantitatives que demande la théorie de la régulation.

Considérons dans le plan A, B deux courbes: $AB = 1$ (hyperbole de Wischnegradski W_1) et la courbe W_2 d'équations paramétriques

$$A = \frac{1 + b^2 - 2ab}{\sqrt[3]{a^2 (1+b^2)^2}},\quad B = \frac{a - 2b}{\sqrt[3]{a\,(1+b^2)}} \tag{4}$$

a, b étant des fonctions du paramètre auxiliaire θ déterminées par les relations

$$e^{-b\theta} = -\cos\theta + b\sin\theta,\quad e^{-(a+b)\theta} = \cos\theta - (a+b)\sin\theta \tag{5}$$

---

* Les constantes sans dimension $A = \delta \sqrt[3]{4T_0/T^2}$ et $B = \frac{T_1}{T}\sqrt[3]{4T_0/T_1}$ figurant dans les équations (3) s'appellent paramètres principaux de la théorie de la régulation directe, ou paramètres de Wischnegradski.

** Remarquons que $x_0 = \frac{\lambda}{\varepsilon}\sqrt[3]{T^2/4T_0^2}$ où λ est la variation relative de la charge. Le cas $x_0 > 0$ correspond à une diminution, le cas $x_0 < 0$ à une augmentation brusques de la charge. Pour étudier le problème des régulateurs dits «à inertie» il aurait fallu, en conservant les équations (3), prendre des conditions initiales plus générales que celles données dans le texte.

*** Ce plan joue ici le rôle d'«espace des paramètres».

Les courbes W_1 et W_2 divisent la portion intéressante ($A > 0$, $B \geqslant 0$) du plan A, B en 3 régions caractérisées par des structures essentiellement différentes, au point de vue de la théorie de la régulation, de l'extension en phase x, y, z en trajectoires (fig. 1) *.

Région des systèmes absolument stables (région *I*), en dehors de W_1. Le segment de repos ** est ici localement stable, et son domaine

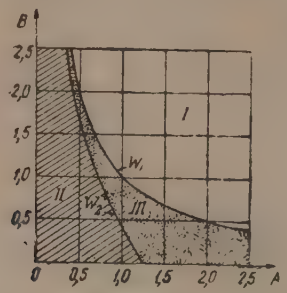

Fig. 1

de stabilité globale embrasse toute l'extension en phase. Le processus de régulation converge quelles que soient les conditions initiales.

Région des systèmes absolument instables (région *II*), entre les axes de coordonnées et W_2. Le segment de repos est localement

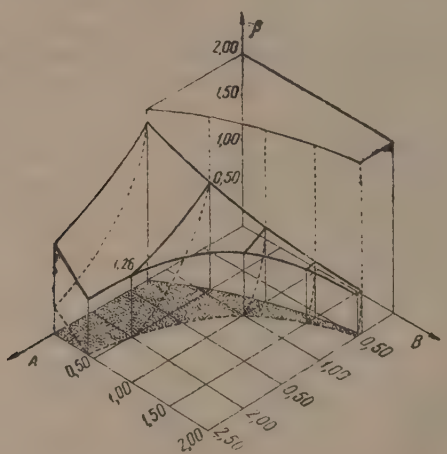

Fig. 2

instable. Le processus de régulation diverge indéfiniment pour toute condition initiale en dehors du segment de repos.

Région des systèmes conditionnellement stables (région *III*), comprise entre W_1 et W_2. Le segment de repos est localement stable, mais son domaine de stabilité globale, limité par les surfaces séparatrices du mouvement périodique instable, n'embrasse qu'une partie de l'extension en phase. Le processus de régulation converge ou diverge selon les conditions initiales.

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* Dans le problème de Mises le plan A, B dégénère et se réduit à l'axe A que la courbe W_1 ne coupe pas et que la courbe W_2 coupe au point $A = 1.2531$ (cf. (6)).

** Nous appelons segment de repos le segment $x = z = 0$, $-1/2 \leqslant y \leqslant +1/2$ dont chaque point est un état d'équilibre du système (3).

3. Dans le cas de stabilité conditionnelle il importe de connaître la décharge admissible maximum, c'est-à-dire la valeur maximum de $|x_0|$ pour laquelle le processus de régulation converge encore vers le segment de repos. Adoptant la représentation de Mises, portons en abscisses le paramètre A, en ordonnés le paramètre B, et en appliquées la grandeur $\beta = 1/|x_0|$, où x_0 caractérise la limite des valeurs initiales, où le processus de régulation converge encore.

Il est facile de donner une limite inférieure de β (très en dessous de la valeur véritable) à l'aide de la condition de Joukowski. D'après Joukowski [7], il faut exiger, en calculant un régulateur centrifuge, que les accélérations initiales, à partir de la seconde transition, soient nulles. On peut montrer que cette condition étant remplie, le processus de régulation, dans le problème de Wischnegradski, converge toujours. Les valeurs de β, admissibles d'après Joukowski, se déterminent facilement à l'aide de l'équation

$$\beta = \sqrt[3]{a(1+b^2)}\, e^{b\theta} |\sin\theta| = \beta(A, B) \qquad (6)$$

où θ est déterminée par la seconde des relations (5), et a, b sont liés à A, B par les relations (4). Ces valeurs de β calculées pour différents A et $B = 0$, 0.5, 1, 1.5, 2, ont été représentées à l'échelle, sur la fig. 2, par des traits forts. Les traits pointillés représentent les valeurs véritables de β. Elles ont été établies approximativement à l'aide d'un procédé graphique proche de celui de Mises, le calcul analytique étant difficile.

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Manuscrit reçu
le 22. XI. 1944.

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Comptes Rendus (Doklady) de l'Académie des Sciences de l'URSS
1945. Volume XLVII, № 5

PHYSICAL CHEMISTRY

PERMUTOID TRANSFORMATION OF FATTY ACID CRYSTALS INTO METALLIC SOAP CRYSTALS AND THE TEMPERATURE DEPENDENCE OF THE TWO-DIMENSIONAL PRESSURE OF MONOLAYERS

By A. A. TRAPEZNIKOV

(Communicated by A. N. Frumkin, Member of the Academy, 26. VIII. 1944)

In earlier communications ([1], [2]) it was shown by the author that crystals of the higher aliphatic alcohols and acids form hydrates, and in the hydrated state adsorb electrolytes from the aqueous phase, markedly influencing the state of the crystals, in particular, their polymorphic transformations and ability to form monolayers. Connected with this is the anomalous, previously described case ([2]) of the kinetics of the formation of a monolayer from a crystal of palmitic acid on a solution of calcium chloride.

The penetration of crystals by electrolytes and the different stages of the interaction of polyvalent cations with fatty acid molecules within the crystal become apparent with particular clearness in a study of the temperature dependence of the two-dimensional equilibrium pressure F_e of the monolayers that form upon spreading from the crystals.

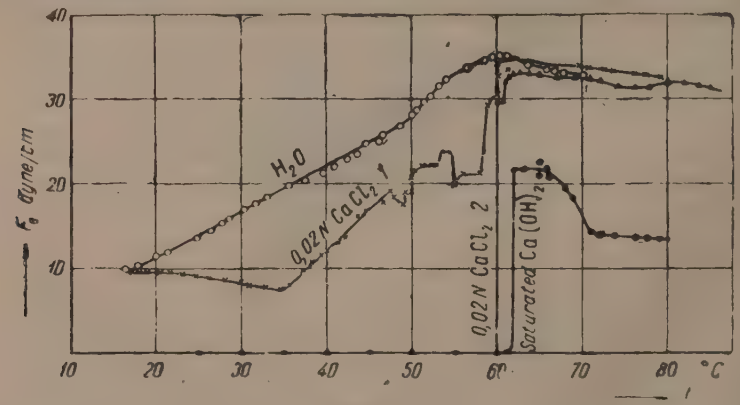

Fig. 1

In Figs. 1 and 2 are given $F_e = f(t)$ curves (obtained during heating) for solutions of 0.02 N $CaCl_2$ (curves *1* and *2* in Fig. 1 are the continuation of the kinetic curves $F = f(\tau)$ in Fig. 1, *a* and *b*([2])), saturated $Ca(OH)_2$, 0.02 N $CuSO_4$, 0.02 N $Th(NO_3)_4$ and pure water, (distilled twice and having a specific conductivity

$$\varkappa = 1.0—1.5 \cdot 10^{-6}\ \Omega^{-1}\text{cm}^{-1}$$

In a broad temperature range, below and above the melting point of the crystals, the influence of electrolytes containing polyvalent cations manifests

itself in the marked lowering of the two-dimensional pressure as compared with the equilibrium pressure on pure water. All along the $F_e = f(t)$ curve there are a number of characteristic temperatures connected with different stages in the transformation and interaction of the crystals with the electrolytes. The most characteristic temperature, which makes it possible to divide the interaction of the fatty acid crystals with polyvalent cations into two essentially different types, I and II, is the melting point, t_m, of the crystals. For palmitic acid in the hydrated state $t_m = 61.5$—62° C, and for the same substance in the anhydrous state $t_m = 62.2°$ C. The two types of intercation, I and II, differ in the magnitude of the pressure and the state of the volume phase of the fatty acid when $t > t_m$.

Type I Interaction. The pressure of a monolayer on the solution of an electrolyte is close to its pressure on pure water. The bulk phase of the fatty acid is in the liquid or anisotropic liquid hydrated state in the form of lenses.

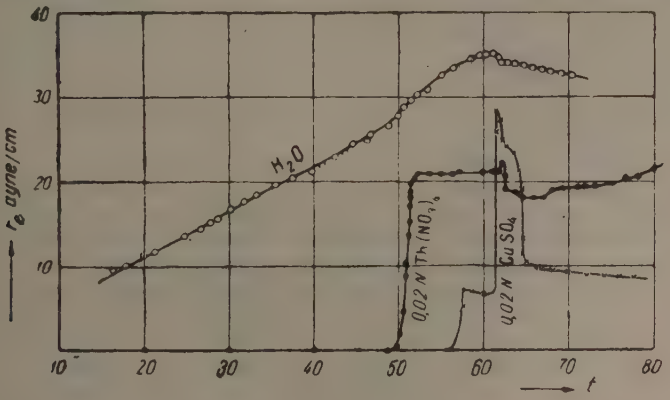

Fig. 2

Type II Interaction. The pressure of the equilibrium monolayer is much lower $\left(\frac{1}{2}\text{ to }\frac{1}{4}\right)$ on an electrolyte solution than on water. The bulk phase of the fatty acid is in the crystal state.

This is corroborated by the data in the accompanying table.

Two-dimensional Pressure of a Monolayer in Equilibrium with Crystals of Palmitic Acid on Different Substances at $t = 75° > t_m$

Type of interaction	Underlying solution	F_e dynes/cm	State of bulk phase	Bulk substance
I	H_2O	31.7	liquid lens	solution of water in palmitic acid
	0.02 *N* $CaCl_2$	33.0	anisotropic-liquid lens	solution of $CaCl_2$ in the liquid hydrate of palmitic acid
II	0.02 *N* $Th(NO_3)_4$	19.5	crystal	a soap—thorium palmitate
	0.02 *N* $CuSO_4$	8.7	»	a soap—copper palmitate
	saturated $Ca(OH)_2$	13.7	»	a soap—calcium palmitate

The crystal state of the volume phase in the type II interaction when $t \gg t_m$ is explained by the fact that we are dealing here not with crystals of the fatty acid but of its metallic soaps. The transformation of a fatty acid into a metallic soap takes place right in the crystal volume phase, as a result of the complete penetration of the crystal of the fatty acid by the solution of the neutral salts of the poly-

valent cation. This process, which may be analysed from the $F_e = f(t)$ curve, proceeds in several stages.

Below a certain temperature a monolayer does not, in general, form from the crystals after the latter have been placed on the surface of the solution (or having formed in the first instant, vanishes, as in the case of the kinetic curve for $0.02\,N\,CaCl_2$ described elsewhere ([2])). The surface tension of the solution σ decreases with the usual temperature coefficient—$\partial\sigma/\partial t = 0.165$ dynes/cm · degree, practically characteristic of pure water.

According to ([2]) this is explained by the slight penetration of the crystal, probably to a depth of several molecular layers, by the polyvalent cations, their adsorption and the formation of a film of a metallic soap. The binding energy of the molecules in it probably is so great that it prevents transition of the molecules from the crystals and the formation of a saturated monolayer. Hence the pressure of the monolayer $F_e = 0$ is not an equilibrium pressure with respect to the internal part of the crystal, consisting of unchanged fatty acid. It is possible that compression of the crystal lattice in the surface layers of the crystal under the influence of adsorbed polyvalent cations binding several molecules of fatty acid, leads to the transformation of the entire crystal into some new, more stable modification.

All along the $F_e = f(t)$ curve below t_m there may be from 1 to 3 temperatures at which it shows sharp discontinuties, owing to rapid increases or even jumps in pressure to certain limiting values. In general, the breaks in the $F_e = f(t)$ curve may be due to changes in the state both of the three-dimensional and the two-dimensional phases, and hence additional experimental data are required for a detailed analysis of the curve. Furthermore, the jumps in pressure are most probably the result of a certain instability in the changes of states. Still it is possible even now to say something about the causes of the kinks in the curve.

For example, in a $0.02\,N$ $Th(NO_3)_4$ solution the rise in pressure at $t = 50°$ is accompanied by visible partial submersion of the crystals of palmitic acid in the water and their partial swelling. This, evidently, is the result of the transformation of the hydrated crystal into another polymorphous modification. With this the outer film of the soap on the crystal bursts, and the molecules of the acid easily come on to the surface of the solution and form a monolayer. With a further increase in t the pressure remains constant until t_m, perhaps also due to the passivating action of the cations, which have been adsorbed on the newly formed modification, and then again makes jumps slightly, which is likewise accompanied by additional swelling of the crystals.

Especially significant is the melting point of the crystals t_m, or a temperature very close to it. In all cases on solutions containing polyvalent cations the pressure jump at t_m when heat is applied, is accompanied by a quite manifest change in the appearance of the crystals: on a $Ca(OH)_2$ solution they lose their lustre; on a $(NO_3)_4$ solution they swell markedly; on a $CuSO_4$ solution they become bright-green. The latter indicates the accumulation of Cu^{++} ions in the bulk of the crystal.

The penetration of the entire crystal by the polyvalent cations becomes possible owing to the sudden loosening of the bonds between the molecules of the fatty acid in the lattice, probably due to the transformation of the crystal into a new polymorphous modification at $t \approx t_m$, which is facilitated by the hydration of the crystal. The decrease in the binding energy of the molecules in the crystal also facilitates their transition and the rapid rise in pressure.

The stratified structure of the lattice of the crystals of the fatty acid and the formation of two-dimensional tunnels between the carboxyl groups adsorbing the electrolyte are advantageous for penetration by the latter. The state of the crystal after penetration and the type of interaction of the fatty acid with the electrolyte are determined by the possibility of a complete substitution of the hydrogen of the carboxyl group by a cation, in accordance with the equation:

$$\nu RCOOH + Me^{\nu+} + A^{\nu-} \rightarrow (RCOO)_\nu Me + \nu H^+ + A^{\nu-} \qquad (1)$$

where ν is the valence of the cation. If the equilibrium constant of this reaction is sufficiently large, substitution already takes place under conditions of comparati-

vely low pH, natural for the given salt solution, as a result of which the crystal of the fatty acid is completely transformed into a crystal of the metallic soap (type II interaction). This is the case in solutions of $CuSO_4$, $Th(NO_3)_4$, and a number of other salts. Thanks to transformation, the crystals do not melt with an increase in temperature even to the boiling point of the solution, since the melting point of the metallic soaps is higher than 100°C. This, as well as analyses of the ash content and a different form of the crystals during observations under the microscope, prove the entire penetration of the crystal by the electrolytes and its complete transformation. In cases of incomplete transformation (in experiments carried out under different conditions) the crystals coagulate to form an amorphous mass (a melt of the fatty acid and its soap) already at $t = 80—90°$ C.

The transformation does not run to completion at the moment that the pressure jumps at t_m. Maintenance of the raised pressure at a temperature 3—4° higher than t_m before it decreases rapidly, is evidently due to the passage of the state of the crystals of the soap through some intermediate stages: either through an intermediate labile modification, from the acid crystals, or through the stage of the loosened lattice under the influence of the not yet squeezed-out ions of the mineral acid, which have formed in the crystal as a result of the exchange reaction (1).

The interaction of the I type, for example, in the case of $CaCl_2$, is limited by the adsorption of the salt on the carboxyl groups (within the crystal) without substitution of hydrogen. This follows on the basis of: 1) the pressure jump to a maximum value corresponding to F_e on pure water, and 2) the visible melting of the crystal with the formation of a turbid lens. The turbidity of the lens, which is retained up to a higher temperature $t > 80°$ than on pure water or on a NaCl solution ([2]), points to an anisotropic liquid state of the hydrate of the fatty acid, due to the orienting action of the Ca^{++} cations (accompanied by Cl^- ions), firmly held by the molecules of the acid even in a melted state. The difference in the type of interaction of a crystal of the acid with Ca^{++} ions in solutions of $CaCl_2$ and $(Ca(OH)_2$ is explained by the difference in pH.

The transformation of the crystals of the fatty acids into crystals of metallic soaps may take place in a number of cases at $t < t_m$ as well; the presence of admixtures, for example, of other fatty acids, markedly lowers the temperature of «bursting» (activization) of the lattice and its penetration by electrolytes. On the other hand; the acid must be in the crystal state; since a drop of the melted acid placed on the surface of the salt solution at $t > t_m$ forms a lens and does not change into a soap. Notwithstanding the thermodynamical advantage of the transformation process, it does not take place for kinetic reasons, since the polar groups of the molecules in the liquid acid are screened by the hydrocarbon chains rolled up into a ball.

The described type of permutoid transformation of crystals of fatty acids into crystals of metallic soaps may take place with the salts of all polyvalent cations at corresponding pH.

On this basis a new method of obtaining metallic soaps has been worked out, which gives exceptionally pure products free, in particular; of univalent cations, invariably present during the production of metallic soaps by the usual method of double decomposition.

The interaction of crystals of fatty acids with electrolytes, discovered in this work, gives additional grounds for affirming that crystals in the hydrated state are three-dimensional analogues of monolayers.The interaction of monolayers with polyvalent cations has been investigated by Langmuir and Schaefer ([3]), and also by the author ([4]).

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Received
26.VIII.1944.

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Comptes Rendus (Doklady) de l'Académie des Sciences de l'URSS
1945. Volume XLVII, № 5

PHYSICAL CHEMISTRY

ON THE TOPOGRAPHY OF THE POLYMERIZATION OF HYDROCARBONS IN EMULSIONS

By A. YURZHENKO and M. KOLECHKOVA

(Communicated by A. E. Arbusov, Member of the Academy, 2. VIII. 1944)

The polymerization of hydrocarbons in emulsions which presents a number of new possibilities (the production of fine dispersions of polymers, co-polymers, *etc.*) and technological advantages, has been the subject of numerous investigations (only part of which unfortunately have been published) ([1–6]). However, only a few of the authors touch upon the mechanism of the process. In each case it is assumed that polymerization takes place in the droplets of the emulsified monomer. Thus, Dogadkin and collaborators ([7]) consider the entire bulk of the monomer droplet acted upon by surface forces as the polymerization zone. Similar views are expressed by Dolgoplosk and others.

This, what might appear to be the most probable, viewpoint is in our opinion essentially contradictory to certain facts. In an ultramicroscopic study of the size of the particles in synthetic (divinyl) latexes, Mikhailova and Bass ([8]) found that their mean radius was 0.05—0.07 mμ. Values in close agreement with these have been obtained with the electron microscope ([9]). Emulsions of the monomer, on the other hand, were found, according to our measurements ([10]), to have a mean radius of 3—5 mμ when common alkaline emulsifying agents were employed. This several thousand-fold expansion of the interface cannot be explained on the basis of the hypothesis that polymerization proceeds in the bulk of the droplet or on its surface. Furthermore, it should be taken into account that the mechanical action applied (agitation in revolving ampules) is very weak, while the gradual transition of the liquid monomer through a viscous fluid to a quasi-solid polymer hampers an increase in the degree of dispersion (especially an increase by one or two orders of magnitude).

The experiments carried out in our laboratory by T. Yurzhenko do not point to any relation between the degree of dispersion (the specific interfacial area) and the rate of polymerization. Our investigation, part of which are described here, permit us to develop a novel conception on the mechanism (topography) of the polymerization of hydrocarbons in emulsions.

We conceive the process of the polymerization of hydrocarbons (rubber producers) in emulsions as proceeding through the stage of the colloidal dissolution of the monomer in the aqueous phase. When 3—5 per cent soap solutions were employed as emulsifying agents, our measurements ([10]) showed that the aqueous phase contains as much as 10—15 p. c. of the monomer. This monomer concentrates mainly in the micellae of the soap (in the zone where the non-polar parts of the molecules of the emulsifying agent accumulate). Growth of the chain takes place in the aqueous phase. This leads to a spontaneous formation of globules of the polymer, which are stabilized against coagulation by the molecules of the emulsifying agent. As the

dissolved monomer is used up, new portions of it pass into the aqueous phase from the hydrocarbon phase. This process, owing to the markedly increased interface area in the emulsion, proceeds very fast and, evidently, cannot limit the process of polymerization. A polymerization process of this nature should lead to the formation of very small particles of the polymer, since their growth is limited by a number of factors. In the case of a weak (albumen) or very dilute emulsifying agent, the forming polymer globules may coagulate into larger aggregates, which leads to the formation of a coagel.

Thus, an emulsifying agent in the first place plays an active (specific) rôle in the polymerization process. This rôle is determined by its hydrotropic action. Other conditions being equal, the polymerization rate should be proportional to the ability of the aqueous phase to dissolve the monomer. This ability is depends upon the nature of the emulsifying agent and its concentration.

In the process of polymerization the emulsifying agent is intensely used up to stabilize the globules of the polymer that are formed ,and the aqueous phase gradually loses its hydrotropic action. Polymerization may sooner or later approach a limit, depending upon the nature and concentration of the emulsifying agent employed.

We carried out several series of experiments on the polymerization of butadiene-1,3 in emulsions with various emulsifying agents. The catalyst employed in most experiments was diazoaminobenzene. The phase ratio was 1 : 1, and the temperature of the experiments, 65° C. The depth of polymerization S was calculated from the volume contraction measured in specially devised vessels with the formula:

$$S = \frac{\Delta v_t}{g\left(\frac{1}{d_m} - \frac{1}{d_p}\right)} 100$$

where Δv_t is the volume contraction of the system during the time t of polymerization; g is the weight of the monomer; and d_m and d_p are the densities of the monomer and polymer, respectively, at the temperature of polymerization.

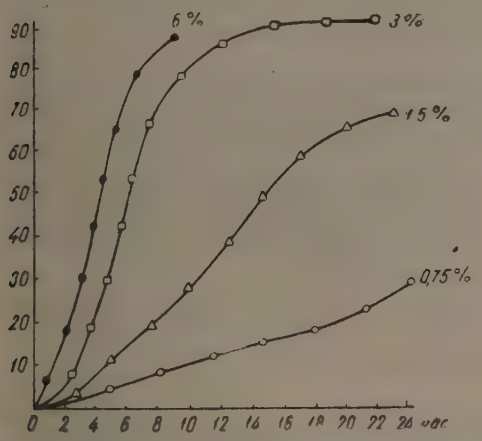

Fig. 1. Polymerization of butadiene-1, 3 at various concentrations of sodium oleate ($c_{Na_2CO_3}$=0.16 g-equiv./liter, c_{DAB} = 2.5 per cent)

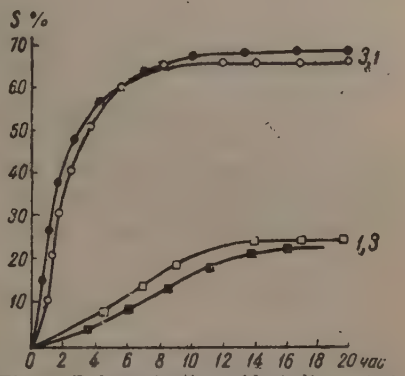

Fig. 2. Polymerization of butadiene -1, 3 at various phase ratios. The aqueous phase contains 3 per cent solution of soduim oleate. The hydrocarbon phase contains 0.3 per cent solution of DAB in divinyl

The following basic data are in agreement with the theory developed:

1. When solid emulsifying agents were employed (kaolin and bentonite suspended according to Tsurinov's method ([11])), polymerization was excluded in all cases, which may be explained by the fact that the aqueous phase is practically not hydrotropic at all.

2. The polymerization rate, when other conditions are equal, markedly increases (frequently nearly linearly) with an increase in the concentration of the emulsifying agent. Moreover, this dependence holds for the concentration range where the surface activity and the emulsifying ability of the solution of the emulsifying agent remain practically constant. The kinetic curves given in Fig. 1 for sodium oleate are similar to those obtained for the other emulsifying agents. The hydrotropic action of the solutions of the emulsifying agents increases with an increase in their concentration in exactly the same manner (see table).

Solubility of Isoprene in Aqueous Solutions of Sodium Oleate (concentration of free sodium carbonate 0.08 g-equiv. /liter)

Concentration of soap in %	$\rho_0 = \frac{n_0^2 - 1}{n_0^2 + 2}$	$\rho = \frac{n^2 - 1}{n^2 + 2}$	$\rho_s = \frac{\rho - \rho_0}{\rho_1 - \rho_0} 100$
0.25	0.2073	0.1075	0.45
0.5	0.2075	0.2079	0.63
1.0	0.2081	0.2085	0.9
1.5	0.2084	0.2098	3.22
3.0	0.2100	0.2142	10.1

3. A series of experiments carried out at various phase ratios (with other conditions—including dispersion—exactly the same) showed that the polymerization rate increases directly as the relative volume of the aqueous phase (Fig. 2). In this case the active rôle played by the aqueous phase as the zone where the polymerization reaction evidently takes place, comes out with especial clearness.

4. If the emulsifying agents investigated be arranged with respect to the rate at which the polymerization proceeds in their solutions, we shall also find them to follow one another in the order of their ability to dissolve the hydrocarbon monomer: namely, sodium (potassium) oleate > ammonium oleate > resin soap > nekal > sodium caseinate.

5. The authors succeeded in obtaining isoprene polymers in the form of latex in soap solutions, when the latter were in contact not with the liquid but the gaseous monomer. Evidently, in this case we were dealing only with a monomer dissolved in the soap.

It is well to point out in conclusion that our concepts cannot be applied to those cases of polymerization of chloroprene in the presence of water described by a number of investigators ([12]), where as a result of polymerization fine dispersions of polymers (latexes) are not formed.

The S. V. Lebedev Plant No. 743.

Received
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Comptes Rendus (Doklady) de l'Académie des Sciences de l'URSS
1945. Volume XLVII, № 5

BIOCHEMISTRY

THE RÔLE OF CARBONIC ACID IN THE FORMATION OF AMINO-ACIDS BY SURVIVING ANIMAL TISSUE

By M. G. KRITZMANN

(Communicated by J. O. Parnas, Member of the Academy, 12. IX. 1944)

The mechanism of the formation of amino-acids from ammonia and α-keto-acids in the animal body and the mode of their formation have not been studied in sufficient detail, this process being generally interpreted as a reversal of oxidizing desamination. Some authors adhere to the hypothesis of «acetylating amination», in analogy with the non-biologic formation of acetyl-alanine from pyruvic acid and ammonium chloride (1). In this reaction the amino-acid formed by condensation of the α-keto-acid and ammonia is hydrogenated and acetylated simultaneously at the expense of a second molecule of pyruvic acid; the first keto-acid may be pyruvic or any other, the second must be pyruvic.

These concepts are based mainly on Erlenmeyer's (1) and Knoop's (2) syntheses, liver perfusions by Embden (3), on Knoop's (4) feeding experiments with α-keto-acids and on *in vitro* experiments of Neber (6) and Krebs (5) with tissue sections.

Neber studied the formation of amino-nitrogen in slices of liver and kidney incubated with pyruvic acid and ammonium salts. Krebs tested also other α-keto-acids, but formation of amino-nitrogen was observed with pyruvic acid and ammonia only. Krebs, therefore, considered pyruvic acid as a specific constituent of the reaction system, and interpreted the mechanism of «reducing amination» in terms of the theory of «acetylating amination». Kaplansky and Karlina(7) have demonstrated the formation of amino-nitrogen, of alanine as determined by the method of von Fürth, and of amino-dicarboxylic nitrogen (Foreman's method) from ammonia with pyruvate or oxalacetate; in experiments with α-keto-glutarate they observed an increase of total amino-nitrogen and of amino-acids dicarboxylic only, while no alanine was formed. In our earlier experiments (8) we observed the formation of amino-nitrogen only in experiments with pyruvic or oxalacetic acids. In the present work with rat liver slices we have found:

1) The formation of NH_2—N from pyruvate and ammonia is rapid in Ringer bicarbonate solution with CO_2, and negligible or nil in phosphate buffer solutions without CO_2; in such buffers the amino—N formation is restored, even after a period of incubation, when bicarbonate is supplied.

2) In oxalacetate solution NH_2—N formaton is rapid even in the absence of bicarbonate and CO_2; with α-keto-glutarate the synthesis is low, but likewise independent of CO_2.

3) Aspartic acid and glutamic acid are formed with pyruvic acid, CO_2 and ammonia, along with alanine, the ratio aspartic acid/alanine being higher than 1 during short incubations and falling to 0 in the second hour.

M e t h o d s. For the experiments tissue slices were used; weighed slices were suspended in Ringer bicarbonate solution in a glass-stoppered flask, aerated with air containing 5 per cent CO_2; in other experiments, the slices were placed in phosphate buffer aerated with CO_2—free air. The substrates were added at the concentrations: pyruvic acid, *m*/10—*m*/20; ammonium carbonate, resp. ammonium phosphate, *m*/50. The flasks were shaken in a water bath at 37—38° C.

In filtrates deproteinized with trichloroacetic acid, estimations were made of amino-nitrogen (after von Slyke), of lactate, malate, alanine and aspartic acid

(by the method of Fromageot and Heitz (9)), and of dicarboxylic amino-nitrogen (Foreman's method, as adapted by Braunstein and Kritzmann).

The data (see table) are calculated as μ-moles per 1 g fresh tissue.

Rat Liver Slices + Pyruvate + Ammonia

Time, min.	$N—NH_2$	Dicarboxylic N NH_2	Lactate	Malate	Alanine	Aspartic acid
			With carbonate			
0	76	16	40	3	5	5
30	156	51	72	15	15	23
60	213	55	90	17	31	33
90	255	59	95	38	60	0
150	368	103	137	63	135	0
			Without carbonate			
0	46	—	51	21	5	8
30	59	—	96	22	12	2
60	59	—	120	24	2	12
90	59	—	144	18	5	10
			m/40 bicarbonate added at 90′			
150	142	—	188	46	77	5
150	72*	—	188*	19*	15*	5*

* Control sample without CO_2, at 150′.

The results prove a specific rôle of carbonic acid, and suggest that the mode of formation of alanine from pyruvate and ammonia in animal tissues is different from the synthesis of alanine (resp. actyl-alanine) *in vitro*. It seems that in liver tissues: 1) the first step is the formation of oxalo-acetic acid from pyruvic acid and CO_2 ($CO_2 + CH_3—CO—COOH \rightarrow COOH—CH_2—CO—COOH$, Wood—Werkman reaction (10)), possibly followed by the production of α-keto-glutaric acid through the tricarboxylic acid cycle. 2) Oxalacetate is aminated to aspartic acid ($COOH—CH_2—CO—COOH + NH_3 + DH_2 \rightarrow COOH—CH_2—CHNH_2—COOH + D$, reductive amination) (and possibly α-keto-glutaric to glutamic acid) by ammonia and unknown hydrogen donors. 3) Alanine is formed by transamination of aspartic acid (resp. glutamic acid) with pyruvic acid ($COOH—CH_2—CHNH_2—COOH + CH_3CO—COOH \rightarrow COOH—CH_2CO—COOH + CH_3CHNH_2—COOH$, transamination). The dicarboxylic amino-acids thus appear as the first products of amination with ammonia, just as they are the central intermediates of transamination. Preliminary experiments with kidney slices testify to a similar mechanism of alanine formation.

This hypothesis on the way of synthesis of amino-nitrogen is supported by further experiments, showing that a synthesis of amino-nitrogen, in the presence of ammonia and pyruvate in bicarbonate free media, can be initiated by the addition of small (catalytic) amounts of dicarboxylic acids subject to transformation into oxalo-acetic acid in the animal tissues.

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Received
5. IX.1944.

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Comptes Rendus (Doklady) de l'Académie des Sciences de l'URSS
1945. Volume XLVII, № 5

GEOLOGY

A NEW STRATIGRAPHICAL SCHEME OF DEVONIAN DEPOSITS OF NORTH-EASTERN KAZAKHSTAN

By N. L. BUBLICHENKO

(Communicated by V. A. Obruchev, Member of the Academy, 15. VII. 1944)

As a result of a many years' work of a large staff of Kazakhstan geologists and of special investigations made by the present author, a great factual material has been gathered on the Devonian deposits, which defies to be forced into any of the existing stratigraphical schemes. Especially abundant new data were furnished within the recent years by the Kazakh Branch of the Academy of Sciences. In this note is suggested a stratigraphical scheme of North-Eastern Kazakhstan compiled in conformity with up-to-date data.

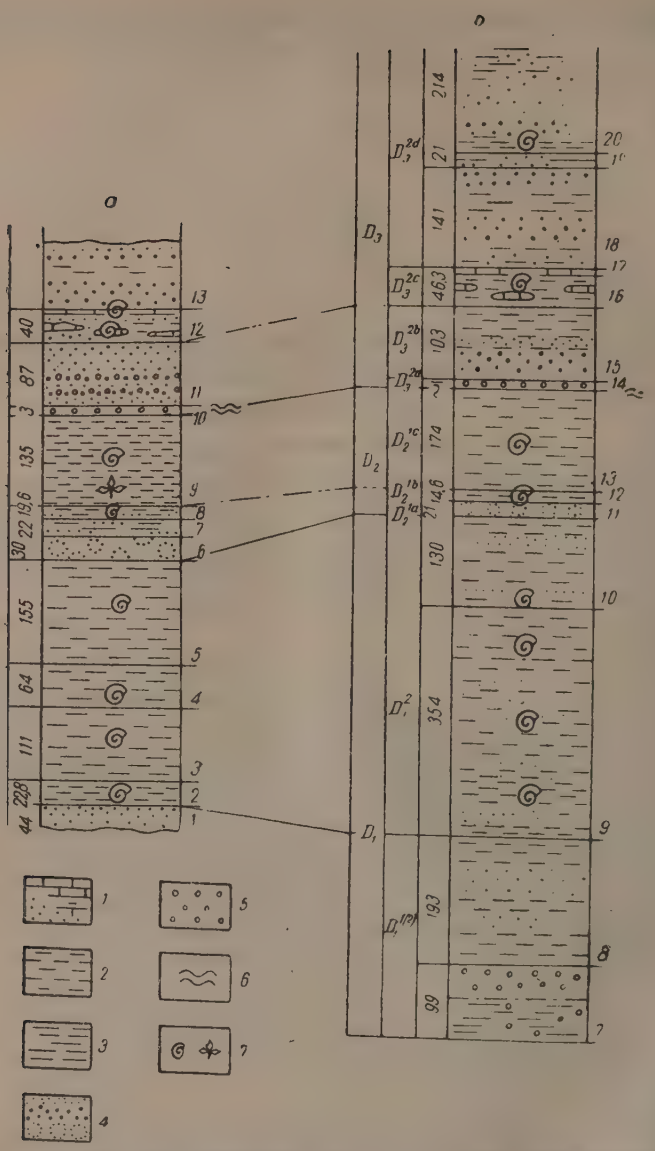
The fauna of the Devonian of North-Eastern Kazakhstan is of a very peculiar, deeply provincial nature, which is especially true of the Lower, Middle and the lower portion of the Upper Devonian

The general habit of many species of North-Eastern Kazakhstan, on the one hand, and of Western Europe and North America, on the other, was found to be much alike although not identical; the vertical range of some identical forms in Kazakhstan seems to be somewhat different from that in Western Europe or North America; a mixing of the West-European and North-American forms is observed; there is a large number of new species. All this made a complete synchronization with the West-European and North-American scales of time impracticable.

In order to bring more precision in the local stratigraphy of so vast a territory as North-Eastern Kazakhstan with its extensive areas of development of Devonian deposits, it will be expedient to work out for it a time-scale of its own; this will permit to obviate the conventionalities in the correlation with the remote subdivisions of Western Europe and, which is most important, will introduce more precision into the subdivision of the strata formations, and make easier the correlation of stratigraphical columns of separate regions of the Kazakhstan zoogeographical province.

The present paper suggests the subdivision of the marine sediments of the Devonian only.

The relations observed between the Silurian and Devonian are transgressive almost all over Kazakhstan. Similar relations were reported from the North-East of the Near-Balkhash area, where erosion of the Silurian deposits had been discovered at the base of the Lower Devonian (Shointas Mts., Kopa). However, in the North-East of the Near-Balkhash region a section (Mt. Kotanbulak) has been found where it is possible to speak of a continuous passage from the Silurian to the Devonian, or a hidden unconformity. Here, between the palaeontologically characterized topmost Upper Silurian and the upper portion of the Lower Devonian, lies a series, nearly 300 m thick, possibly referring to the Gedinnian stage. To the Gedinnian is also probably referable the fauna of the upper course of the Sokur River (south-eastward of the Karaganda basin), with *Leptostrophia rotunda* sp. n. It should be mentioned, however, that no reliable palaeontologically characterized strata of the Gedinnian stage have so far been found in Kazakhstan.



Region of the Ainabulak well (Kotanbulak Mts.), N.-E. Near-Balkhash area. *a*, stratigraphic column of the N.-E. tectonic block located N.-E. of the well; *b*, stratigraphic column of the northern tectonic block located north of the same well. Thickness is given in metres. Legend: *1*, limestones; *2*, aleurites; *3*, clayey shales; *4*, common sandstones and sandstones with a ball parting; *5*, conglomerates; *6*, transgressive occurrence; *7*, fauna and flora

D_1^2. The work of the recent years has furnished a very abundant fauna of the upper part of the Lower Devonian (Sardjalian stage) provisionally correlated with the Koblentz stage of Western Europe. A determination of the Sardjalian stage was for the first time given by the present author.

In this paper the author makes an attempt to bring more precision into this conception on the basis of a much more abundant material than the one previously available. Without renouncing to the old name given after the Sardjal Mountains located East of the town of Karkaralinsk, I will but change the type of this stage, choosing series 2, 3, 4 and 5 (figure, *a*) of the stratigraphical column of the North-

Eastern tectonic block located East of the Ainabulak well (Kotanbulak Mts., North-Eastern Near-Balkhash region).

The Sardjalian stage is determined by the presence of *Acrospirifer primaevus* Stein., as well as by the appearance in the lower part of *Leptostrophia beckii* Hall., and for the given section of the Kotanbulak Mts., of smooth pelecypods; in the upper portion the Sardjalian stage terminates with the appearance of *Proschizophoria magnifica* sp. n. A general list of the principle forms encountered in the Sardjalian stage is as follows: *Parmorthis triangularis* Zeil. *, *Leptostrophia explanata* Sow.*, *L. beckii* Hall., *L. sowerbyi* Barr.*, *Stropheodonta ampla* sp. n. *, *S. ampla* sp. n.*, *S. bella* sp. n., *S. virgata* Drev., *S. piligera* Sandb., *Leptaena bokéi* Barr.*, *Leptocoelia biconvexa* sp. n.*, *Chonetes granda* sp. n., *Camarotoechia* ex gr. *daleidensis* Roem., *Acrospirifer primaevus* Stein.*, *A. rousseau* M. Rouault, *A. cabedanus* A. et V., *A. arduenensis* Schnur., *A. assimilis* Fuchs., *A. schointass* sp. n., *A. mediobalchaschensis* sp. n., *Schizospirifer predivaricatus* sp. n., *Sch.* aff. *daleidensis* Stein., *Sch. solitarius* Krantz.*, *Sch.* sp.*, *Cyrtina heteroclita* Defr., *Nucleospira maillieuxi* Dahm.*, *Pelecypoda* ind., *Conularia* sp., *Gastropoda* ind.

Of the vegetable remains *Blasaria sibirica* Zal. has been determined. The thickness of the Sardjalian stage in the Kotanbulak makes up 484 m.

Aside from the North-Eastern Near-Balkhash region, a faunistically characterized Sardjalian stage has been established in the region of the Bosaga station, South-East of Karaganda, West of Karkaralinsk and near Bayanaul.

The D_2^1 strata of the Kazakh stage quite conformably supersede the Sardjalian stage. The boundaries of the former are determined in the lower part by the appearance of the first *Proschizophoria magnifica* sp. n.; above, by the disappearance of *Acrospirifer cotanbulack* sp. n., and in the considered sections of Kotanbulak Mts., by the transgressively overlying Famennian stage. The 11, 12, 13 series (figure, *b*) of the column of the northern tectonic block of the Kotanbulak Mts. have been chosen as the type of the Kazakh stage; indeed, it is only the stratigraphic position and the resemblance of the general type of fauna that made it possible to correlate them with the Eifelian or, according to the most modern nomenclature, with the Cauvain stage of Western Europe.

A general list of the principal forms of Kazakh stage is as follows: *Rugosa* ind., *Receptaculites* sp., *Pleurodictyum* sp., *Semicosciniu m* cf. *striatum* Nekhor., *S.* sp., *Hemitrypa* sp., *Dalmanella berkara* sp. n.*, *Parmorthis* cf. *triangularis* Zeil., *P.* aff. *tetragona* Vern., *Proschizophoria magnifica* sp. n.*, *Schuchertella umbraculum* Schloth. *, *Leptaena rhomboidalis* Wilck.*, *Stropheodonta* aff. *sowerbyi* Barr., *S.* cf. *bella* sp. n. *S.*, cf. *patersoni* Hall., *S.* ex gr. *naranjoana* Quenst., *Chonetes sarcinulata* Schloth.*. *Chonetes planulata* Stein. sp. n., *Ch.* aff. *coronata* Conr., *Productella spinulicosta* Hall., *Eatonia* sp., *Camarotoechia meganteris* sp. n.*, *Atrypa reticularis* Linne*, *Leptocoelia biconvexa* sp. n.*, *Acrospirifer cotanbulack* sp. n.*, *A. juschi* Spriesterbach*, *A.* (?) *pseudohystericus* sp. n., *Schizospirifer divaricatus* Hall., *Mediospirifer* (?) aff. *mucronatus* Conr., *Spiriferellina* aff. *subsulcata* Barrois, *Pelecypoda* ind.*, *Platyceras* sp.*, *Pleurotomaria* sp., *Orthoceras* sp., *Nautiloidea* ind., *Aphyllites* sp., *Pelecypoda* ind.*, *Agoniatites bohemicus* Barr., *A. kayseri* Wdkd., *Anarcestes crispus* Barr., *Clarceoceras umbonale* Wdkd., *Trilobita* ind., *Pisces* ind.

Of the vegetable remains *Blasaria sibirica* Zal. has been determined. The total thickness of the Kazakh stage in the region of the Kotanbulack Mts. is 151—219.7 m.

In this region the strata of the Kazakh stage are transgressively overlain, without apparent angular unconformity, by strata of the Famennian stage.

D^2_2. As the type of the Aidarly stage (the upper portion of the Middle Devonian), the calcareo-areno-marlaceous strata occurring along the route Bayanaul—Pavlodar, South of the Mt. Aidarly, have been selected.

The list of forms encountered in the Aidarly stage is as follows: *Crypophyllum* sp., *Lythophyllum corneolum* Wdkd., *Hemitrypa* cf. *bugusinica* Nekh., *Kayserella*

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* In this list, as well as in all the subsequent ones, the forms which have been found in layers accepted as the type of a given stage are marked with an asterisk (*).

raricosta sp. n., *Schuchertella umbraculum* Schloth., *Strophcodonta taeniolata* Sandb., *S. interstrialis* Phil., *Chonetes scitulus* Hall., *Productella spinulicosta* Hall., *Uncinulus grandis* sp. n., *Atrypa reticularis* L., *Acrospirifer undiferrus* F. Roemer, *Schizospirifer* cf. *divaricatus* Hall., *Mediospirifer macronotus* Hall., *Athyris concentrica* L. v. Buch., *Cryptonella planirostra* Hall., *Pelecypoda* ind., *Gastropoda* ind.

The Aidarly stage has been encountered in the sections of the North-Eastern Near-Balkhash region with a thickness up to 396 m; these strata have an extensive development also in the other regions of North-Eastern Kazakhstan.

D^1_3—Maian layers (or Maian stage, after the author's nomenclature) have been established by D. V. Nalivkin(1) and named after Lake Maikuben, Bayanaul region. Nalivkin does not mention the typical occurrence, and therefore the author has selected as such the exposure of these layers near the Akkozu Mts. (East of Mt. Aidarly), where they are clearly seen to lie above the Aidarly stage.

The chief forms are as follows: *Rugosa* ind., *Bryozoa* ind., *Dalmanella* sp.*, *Productella subaculeata* Murch., *P. spinulicosta* Hall., *P. proarctirostrata* sp. n.*, *P. suspecta* sp. n.*, *Camarotoechia sapho* Hall., *Atrypa reticularis* Linné, *Acrospirifer undiferoides* sp. n., *Mediospirifer ali* Nal.*, *M. seid* Nal.*, *Cyrtospirifer* (?) *achmet* Nal.*, *Athyris spiriferoides* Eaton, *Pelecypoda* ind., *Trilobita* ind.

Nalivkin (1) assigns the Maian layers to transitionary Middle-Upper Devonian deposits; the present author believes it quite possible to refer them to the base of the Upper Devonian. From the data so far available, the upper and the lower boundary of the stage are determined by the appearance or the disappearance of a very characteristic form—*Mediospirifer ali* Nal. In Kazakhstan the Maian stage is rather extensively developed, but it always retains its provincial character.

D^2_3. Out of the marine deposits the Famennian is the most widespread in North-Eastern Kazakhstan and is rather persistent. At many places a transgressive occurrence on the underlying layers has been recorded (as, *e. g.*, the occurrence on the Kazakh stage in the Kotanbulak Mts., North Eastern Near-Balkhash region, *etc.*).

The fauna of the Famennian stage has the least provincial habit, although its peculiar nature is beyond any doubt. Even the guide form, *Cyrtospirifer sulcifer* H. et C., may possibly be not identical, but only a vicar North-American form.

Nalivkin subdivides the Famennian stage into two parts: the lower, the Meister layers (guide form *Productus meisteri* Peetz), and the upper, the *Sulcifer* layers (guide form *Cyrtospirifer sulcifer* H. et C.).

A similar subdivision is observed in a number of places in Kazakhstan; it should be noted, however, that it is borne out by the lithological composition of the rocks containing the fauna: the Meister layers are usually composed of pure limestones—in this case the fauna has also a typical habit of the Meister layers. Accordingly, the calcareo-areno-marlaceous composition predetermines the typical faunistic character of the *Sulcifer* layers. On the other hand, when within the lower layers of the Famennian stage pure limestones are absent, the complex of fauna of the Meister layers is absent, too. The question as to the mutual relations of the Meister and *Sulcifer* layers cannot be considered as finally settled.

The deposits of the Famennian stage are gradually superseded by Lower Carboniferous ones; it is only in the Djezkazgan region and possibly in the North-West of the Near-Balkhash area that some elements of a transgressive occurrence are observed.

In the present paper the extensive development of the layers with a fauna of the Koblentz type of the upper part of the Lower Devonian, and the fauna of the lower portion of the Middle Devonian are for the first time treated in some detail; greater precision is also introduced into the stratigraphy of the Upper Devonian.

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Received
15. VII. 1944.

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Comptes Rendus (Doklady) de l'Académie des Sciences de l'URSS
1945. Volume XLVII, № 5

GEOLOGY

SHOR-KAN, A SULPHATE DEPOSIT OF A NEW TYPE IN THE SOVIET UNION

By K. M. FEODOTIEV

(Communicated by D. S. Beliankin, Member of the Academy, 17. VI. 1944)

The Shor-kan deposit of mirabilite was discovered by S. Mukimov in 1940, in the complex of the salt deposit of the Kamysh-kurgan salt region, Tadjik SSR. In 1942—43 it was investigated by the present author to whom it appeared of great interest because of its having been formed by deposition of salts rocks of the «gypsiferous series» underlying the Shor-kan deposit.

This note contains a brief geological characteristic of the deposit and a description of the sulphate occurrence and of the brines from which it originated.

The lowland parts of the Asht region, Leninabad province, where the Kamysh-kurgan salt region is situated, are composed of saline rocks of Tertiary and Quaternary age. The oldest among them are clays alternating with rock-salt beds, forming a complex of rocks of the so-called «saliferous series» ([1],[5],[8]). This complex is overlain by a series of clays, sandstones and gypsums, the so-called «gypsiferous series». In various parts of the region this series has been found to contain tenardite intercalations. There is good reason to believe that it is from here that are removed the sulphate salts which in dissolved condition are found in the Shor-kan brines. We will therefore dwell here at some length on this series.

The «gypsiferous series» in its lower parts is composed of clay beds, green or brownish-green in colour. These clays are highly schistose; they interbed with gypsum which is represented by intercalations or beds of considerable thickness, but highly contaminated with the same clay. In the upper portions of the series, proceeding upward the section, sandstones make their appearance. They become more abundant in the upmost strata and finally the clays are entirely superseded by them.

In the exposures of the gypsiferous series, near the Shor-kan deposit (Makhau-tau Range) there is found scarcely any rock save for sandstone. Their beds are of a reddish-pale yellow colour of different shades; they are interbedding with thin gypsum layers. In the Makhau-tau Range small tenardite intercalations are also observed. At the Makhau-tau Range the total thickness of this series attains 1300—1350 m ([1]). The gypsiferous series of rocks is overlain by Quaternary sandstones, clays, and conglomerates. They contain also intercalations of marls and gypsum, but are distinguished by a lower salt concentration.

Nearly all of the above-listed rocks, save for the youngest, have been crumpled into arch-like folds by the tectonic processes of Alpine folding. These folds have a general strike from NW to SE, with their convexity turned to the NW. Within the Kamysh-kurgan salt region they extend over a distance of about 70 km. The extreme of these, the south-eastern one, is the Makhau-tau Range, represented by the gypsiferous series; it participates directly in the formation of the Shor-kan deposit;

it is on the extension of the north-western limb of this anticline, the Djanbulakchon Ridge, that the deposit is located; the supply of material to the deposit is connected with the redeposition of soluble sulphates of the gypsiferous series of Makhautau.

The surface of the Shor-kan deposit presents a saline, covered with a crust of salts, developed over an area of 0.3 km². The saline area is stretched from SW to NE. Its configuration is of a rather queer shape, repeating the outlines of the depression.

A reconnaissance of the saline showed that following its long axis is situated a mirabilite occurrence of a undersand type; it is of a considerable thickness and is stretched from SW to NE, its shape being that of a flattened fusiform body. Its narrow portion, about 200 m long, about 20 m across, and 3—4 m thick, occupies the north-eastern area. The mirabilite rock occurring here may be characterized by the data of drill-hole 18, given in Table 1.

Table 1

Chemical Characteristics of Mirabilite Occurrence

No. of holes	Depth of sampling, m	Thickness of strata sampled, m	Chemical composition of dried sample in weight %									Weight %	
			Insoluble residue	Na^+	Ca^{++}	Mg^{++}	Cl'	SO_4''	HCO_3'	H_2O	Total	Na_2SO_4	NaCl
18	0.5—1.0	0.5	25.06	19.66	3.57	0.15	5.03	42.71	0.91	2.38	99.47	52.12	8.54
	1.0—1.5	0.5	17.08	12.96	3.93	0.34	4.10	31.56	0.91	29.64	100.52	44.84	9.52
	1.5—2.0	0.5	20.78	14.05	3.71	0.28	3.54	34.07	0.61	22.94	99.98	47.12	7.56
	2.0—2.5	0.5	16.80	13.89	5.00	0.32	4.47	35.14	0.91	23.66	100.19	44.34	9.62
	2.5—3.0	0.5	12.12	13.50	2.28	0.24	4.10	28.35	0.91	38.10	99.60	47.53	10.93
	3.0—3.5	0.5	14.04	18.88	3.64	0.19	6.24	39.83	0.91	16.56	100.29	54.76	12.28
5	0.0—0.5	0.5	60.82	3.19	4.28	0.24	1.48	14.69	0.91	14.44	100.05	31.76	6.83
	1.0—1.5	0.5	9.28	20.30	1.64	0.10	2.03	44.15	0.61	21.86	99.97	75.13	4.26
	1.5—2.0	0.5	19.72	18.53	3.28	0.19	2.60	43.16	0.91	12.36	100.75	64.24	5.27
	2.5—3.0	0.5	11.74	21.78	2.71	0.19	1.11	50.86	0.61	12.00	100.70	73.11	2.24
	3.0—3.5	0.5	2.36	14.67	1.14	0.15	1.11	32.10	0.61	47.62	99.76	82.74	3.49
	4,0—4.5	0.5	16.97	24.75	1.64	0.21	1.73	32.09	0.91	1.22	100.34	73.63	2.84
2-IV	1.5—2.0	0.5	33.96	14.45	4.79	0.37	4.58	36.13	0.31	5.42	100.01	37.47	7.98
	2.0—2.5	0.5	42.20	10.68	4.00	0.28	6.24	32.50	0.31	4.60	100.81	21.28	10.44
	2.5—3.0	0.5	36.52	13.39	5.15	0.24	6.45	35.68	0.31	3.02	100.76	34.10	10.88
	3.0—3.5	0.5	29.38	18.77	2.64	0.04	2.08	42.21	0.91	3.72	99.75	56.04	8.56
	3.5—4.0	0.5	33.56	15.23	4.22	0.26	4.16	37.39	0.31	4.46	99.59	40.69	7.20

The central portion—200 m along the axes, 175 m wide, up to 5.5 m thick—is a swell filled with the purest and richest rock, which in the same table is characterized by hole 5 corresponding to the purest areas, and by hole 2-IV corresponding to the impure portions of the marginal zone.

These data, as well as the materials concerning the other holes (a total of 32 holes were drilled), which for lack of space are not given here, show that in some parts the mirabilite rock is highly contaminated with clay, which my be seen from the high values of the insoluble residue and the large amount of sodium chloride.

Hypsometrically, the deposit lies above the lowest part of the Kamysh-kurgan valley. At the lowermost point wedge out the fresh waters flowing down the Kurala Range, which supply the villages situated along the line of their issuance.

These fresh waters seem also to have taken part in the formation of the Shor-kan deposit. They percolate to a considerable depth following small tectonical disturbances («gypsum tectonics») (4), dissolve the sulphate deposits syngenetic with the gypsiferous series, and, coming out to the surface saturated with sodium sulphate, precipitate mirabilite upon cooling. Mineral springs are found at the deposit itself.

The chemical composition of the brines from drill-holes is given in Table 2. Analysis No. 216 corresponds to the brine from hole 5-IV drilled near a spring; the

two other analyses, to the brines from holes sunk at a distance of 200—400 m from the spring.

Table 2

Chemical Composition of Brines from Shor-kan Deposit*

		No. 216, hole 5-IV, 15. IX. 1942		No. 412, hole 9-IV, 19. X. 1942		No. 267, hole 7-II, 17. IX. 1942	
		g/l	mg-equiv.	g/l	mg-equiv.	g/l	mg-equiv.
Specific gravity at 20°C		1.195		1.195		1.227	
Dense residue		291.600	—	285.834	—	321.200	—
Ionic composition	$Na^{\cdot}$	99.284	4318.99	97.408	4221.84	110.500	4803.00
	$Ca^{\cdot\cdot}$	1.429	71.307	2.877	114.121	2.858	142.614
	$Mg^{\cdot\cdot}$	1.747	143.603	2.359	193.910	1.966	161.605
	SO_4''	107.607	2235.22	89.090	1853.07	90.119	1874.47
	Cl'	81.132	2287.92	94.890	2675.89	114.512	3229.23
	HCO_3'	0.458	0.751	0.305	0.500	0.305	0.500
Salt composition	Na_2SO_4	143.85	—	109.82	—	111.33	—
	NaCl	133.95	—	156.87	—	188.78	—
	$MgSO_4$	8.55	—	11.68	—	9.73	—
	$CaSO_4$	4.81	—	8.32	—	9.67	—
	$Ca(HCO_3)_2$	0.47	—	0.31	—	0.31	—
Total of salts		291.63	—	287.00	—	319.82	—

*Sampling by K. M. Feodotiev; T. I. Sukonkina, analyst.

The large analytical material available shows that the composition of brines is relatively constant both within the whole area of the deposit and during different seasons. The temperature of the brines is constant throughout the year; it varies within a narrow interval by about +17° C.

The figurative points of the composition of brines recalculated after Eneke ([6],[7]) and plotted on the diagrams of mutual solubility of salts in the system $Na_2Cl_2 + MgSO_4 \rightleftarrows Na_2SO_4 + MgCl_2$ ([2],[3]), show that at +25° the brines fall into the field of tenardite, and at 0°, into the field of mirabilite. It follows therefore that upon cooling mirabilite is deposited from the brines, while upon evaporation in summer, it is tenardite, which fact is supported by observational data.

This type of natural brines has never been encountered before.

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Received
17. VI. 1944.

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Comptes Rendus (Doklady) de l'Académie des Sciences de l'URSS
1945. Volume XLVII, № 5

GENETICS

HETEROCHROMATIZATION AS A CHANGE IN THE CHROMOSOME CYCLE

By A. A. PROKOFYEVA-BELGOVSKAYA

(Communicated by A. A. Richter, Member of the Academy, 13. VII. 1944)

The study of chromosomes in the cell nuclei of the salivary glands of *Drosophila melanogaster* in the mosaic lines w^{mt},sc^{8}, w^{mt4},w^{mt5}, rst^{3}, f^{27}, f^{59},f^{168} has revealed regular changes in the active sections containing the loci listed above, approaching the inert region ([1], [2]). These changes as a whole—a reduced conjugation of chromonemes and an increase in the content of thymonucleic acid in this section, are characteristic of the state of the inert regions of chromosomes in the cell nuclei of salivary glands. In the nuclei of ganglial cells these regions tend towards heteropicnosis ([3],[4],[5]). We believe that the changes described in chromosomes as heterochromatization and heteropicnosis are a change of state a chromosome undergoes in the cell cycle, and that these changes are but a premature transition of the chromosome's definite section to the «mitotic» condition. The separate section of a chromosome go through their cycle in a rather independent way, displaying heterochromatization while the main complex of the nucleus is in the state of rest.

In the process of evolution the chromosome has become differentiated into inert and active regions of different biochemical structure ([6]), responding in a different way to changes in developmental conditions ([7],[8]), and behaving in a different way in the cycle of nuclei ([9],[10]).

The active regions are thoroughly specialized, stable sections of the chromosome, whose response to varying external or intracellular conditions is a relatively weak one. The inert region are the most labile sections of the chromosome; their specialization is low, they are sensitive to the changes in genetical, developmental and environmental conditions, as well as to the physiological changes that take place in the tissues (see also ([7],[8])). The inert regions are the first to respond to any change in the development conditions by variations in their cycle. Active regions that are located in the proximity of the inert ones are involved in the sphere of the latters' activity, and similarly undergo reversible changes of their cycle, physiologically and cytologically transforming into «inert» regions, although being fully heterochromatic. Thus, actively responding to the changes in genetical, developmental and environmental conditions by changing their cycle (reducing or extending the metabolic stage), the inert regions are acting upon the adjoining active regions which participate directly in the organism's development, affecting thereby the course of development.

The duration of the metabolic stage of the inert regions and of the adjoining active regions varies as to the cell. In other words, the period of action of the given active regions in a metabolic nucleus is entirely dependent on their cycle which depends on the cycle of the inert regions located in their proximity.

In mosaic lines of *Drosophila melanogaster* (w^{mt4},w^{mt5},rst^{3}, sc^{8},f^{B27},f^{B28},f^{B168},*etc.*) all these cellular processes bring about strictly definite developmental alterations:

Hereditary Determination of the Chromosome Cycle (of the 1AB1 (*y-ac*) 20 ABC region)

Stock	Structure of individuals	Number of nuclei investigated	State of the 1AB1-2 (*y-ac*) 20 ABC region in the sc^8 chromosome, %		
			heterochromatic	euchromatic	heterochromatic
♀ $sc^8 \times$ ♂ $\widehat{xy^{sh}}$	$\frac{sc^8}{\widehat{xy^{sh}}}$	200	13	87	—
♀ $sc^8 \times$ ♂ sc^8	$\frac{sc^8}{Y}$	200	16	84	—
♀ $sc^8 \times$ ♂ *yacv*	$\frac{sc^8}{yacv}$	200	20	80	—
♀ $sc^8 \times$ ♂ sc^8	$\frac{sc^8}{sc^8}$	200	39	34	27
♀ *yacv* × ♂ sc^8	$\frac{yacv}{sc^8}$	200	71	29	—

1) a curtailed period of the operation of a given active region in the metabolic nucleus affects the trend of development of a given character towards its recessive manifestation; 2) the variability of the cycle of a given chromosome region in different cells, being due to the sensitivity of the inert regions, is responsible for the mosaic nature of the manifestation of the character itself; 3) the degree of the variability («heterochromatization ratio», see above) is constant for each given line. Thus, according to our interpretation, the trend of development of a character in mosaic lines of *Drosophila* stands in relation to the length of the stage of «metabolic nucleus» in a given chromosome region.

A study of the heterochromatization of the section *y-ac* in the sc^8 chromosomes in nuclei of the salivary glands of individuals of a $\frac{sc^8}{\widehat{xy^{sh}}}$, $\frac{sc^8}{Y}$, $\frac{sc^8}{yacv}$, $\frac{sc^8}{sc^8}$, $\frac{yacv}{sc^8}$ constitution revealed in them a different heterochromatization percentage, constant for each of these categories of flies. Analysing these results in relation to the cycle of the chromosome; we find that the duration of the metabolic stage of the *y-ac* section in the cs^8 chromosome is the shortest in $\frac{yacv}{sc^8}$ females, while it is the longest in $\frac{sc^8}{\widehat{xy^{sh}}}$ females. The duration of the metabolic stage of the *y-ac* region in the sc^8 chromosome for all cases investigated may, according to the heterochromatization percentage; be expressed by the following series:

$$\frac{sc^8}{\widehat{xy^{sh}}} > \frac{sc^8}{Y} > \frac{sc^8}{yacv} > \frac{sc^8}{sc^8} > \frac{yacv}{sc^8}$$

This series strictly corresponds to the types of mosaicism found in the respective fly categories by Nuzhdin ([15],[16]).

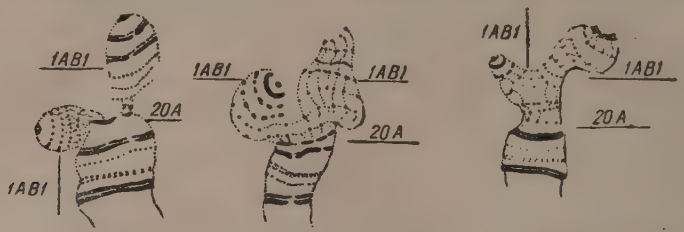
We thus are brought to a conclusion that the hereditarily conditioned degree of mosaicism of given fly categories is due to the hereditary character of the cycle of their *y-ac* section.

If we were to consider these facts in the light of the interpretation given by us to the cycle of chromosome regions, the following picture would be obtained: it is not without import for the cycle of a given chromosome region whether during meiosis of the parents this region was in a hetero- or homozygous condition. The section *y-ac* that passed through meiosis of the parents in a homozygous condition (sc^8 chromosome supplied from the mother in a cross ♀ $sc^8 \times$ ♂ *yacv*) has a longer metabolic stage in the progeny.

The section *y-ac* that passed through meiosis in a heterozygous condition, being in the condition of heteropicnosis (a sc^8 chromosome supplied from the father

* *yacv*-chromosome attached to the short arm of the Y.

in a cross ♀ *yac* × ♂ *sc*[8]), has a shorter metabolic stage in the progeny. This is why we observe differences in the structure of homologous chromosomes in a *sc*[8] female in the most sensitive inert region (20 ABC) and in the adjoining active region (1AB1) with the loci *y-ac*: in the one chromosome (supplied from the father) the given region has been heterochromatized, while in the other it is in an euchromatic condition (see figure). This cytological evidence goes fully to confirm the correct-



The 1AB1—20 ABC region of the *sc*[8] chromosome in female larvae

ness of Nuzhdin's conclusion as to the genetic heterogeneity of chromosomes of homozygous females in the pure line *sc* ([17]).

We thus are brought to conclude that in the case investigated the duration of the metabolic stage of the *y-ac* region in the *sc*[8] chromosome is hereditarily determined by three agents:

1. The position of a given chromosome section relative to the inert region.
2. The condition in which a given chromosome region had passed through meiosis in the parents.
3. The presence in the nucleus of a system of «modifiers» in the form of additional inert regions (*Y*-chromosome).

It seems probable that all those agents underlying the hereditary nature of the chromosome region cycles are of general significance (table).

In their study of the locus white Belgovsky and the present author have shown([11]) that mutational changes of this locus are attended by heterochromatization and changes in the cross-over properties of the respective chromosome region (3C2). Our observations on the structure of regions of the X-chromosome involving the loci Hairy-wing (1B1-2), Bar (16 A), bobbed (20 C) show that the cases of minute rearrangement are also accompanied by a state change of these regions ([12],[13]). There is evidence to the effect that changes in the cycle of a chromosome region in mosaic lines of *Drosophila* are accompanied by mutational alterations known for the given locus ([14]).

From an analysis of this material we may infer that not only changes in chromosome regions, in case these are brought near to an inert region, but other types of inherited variation as well, *e. g.* point mutations and minute rearrangements, are followed by local hereditary changes in the cycles of the given chromosome regions. Some of the recessive mutations may be due to the reduced metabolic stage in the given loci. The correctness of this assumption is supported by a vast material on hypomorphous and amorphous mutations ([17]).

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Comptes Rendus (Doklady) de l'Académie des Sciences de l'URSS
1945. Volume XLVII, № 5

PLANT PHYSIOLOGY

ACCUMULATION OF RUBBER BY KOK-SAGHYZ AT DIFFERENT PERIODS OF ITS LIFE

By A. A. DROBKOV

(Communicated by D. N. Prianishnikov, Member of the Academy, 18. VI. 1944)

It has been noted elsewhere that, notwithstanding its theoretical and practical importance, the problem on the formation of rubber at different periods of the life of kok-saghyz remains as yet unexplored. A number of writers believe that the highest yield of roots and of rubber per unit area is obtained from one-season-old, normally developed kok-saghyz plants towards the end of the growing season; on the second year, according to this point of view, the total yield of rubber remains the same, owing to the death of part of the plants during the summer, although the rubber content in individual roots becomes higher. For this reason it is advised to harvest kok-saghyz in the autumn of the first year, and leave for the second year those plantations only which remained for some ground underdeveloped ([2—4]). Other students consider it to be more advisable to cultivate kok-saghyz during two seasons and to harvest it on the second year at the end of July, for, according to their studies, the roots of kok-saghyz contain the largest amount of rubber just at this time.

Since 1941 we have conducted a series of studies on the accumulation of rubber in the roots of kok-saghyz as connected with the fertilizers supplied, the period of life and other factors. In the present paper only a part of the problems which were studied by us at the Kazan State Plant-Breeding Station is dealt with.

In the table below are presented the results of the field experiments on the effect of fertilizers, and of the time of harvest; kok-saghyz was harvested on three dates.

The first period of harvest: one-season-old plants, harvested on the 6th of November, 1941. The rubber content ranged from 3.79 to 5.78 per cent, according to the fertilizer. Without going into detail, we shall only note that the rubber content was low at this period; its average value was 4.5 per cent. The other date of harvest was the 5th of May, 1942. Towards this time, as a rule, the rubber content exceeded more than twice the above figure, irrespective of the fertilizer used; it reached, namely, an average of 11.89 per cent. The third date of harvest was the 10th July, at the end of fructification. The rubber content was but slightly changed as compared with the spring; on the average it was equal to 11.96 per cent.

The figures in the first graph show how many kilograms of nitrogen, P_2O_5 or K_2O were distributed per 1 hectare. The soil was medium podzolized clayey soil.

In an other experiment started in the spring 1942 at the same station (the soil being medium podzolized clayey soil again) we were dealing with the problem as to the changes occurring in the rubber during the winter, under the influence of low temperatures, in the spring on the second year of the plant's life, before the beginning of the growing season, and at the later periods. The first samples for the deter-

Rubber Content of Kok-saghyz as Dependent on Fertilizer and Time of Harvest

Scheme	Rubber content in per cent to dry weight		
	Harvest on 6.XI	Harvest on 5.V	Harvest on 10.VII
No fertilizers . .	3.97	10.35	10.46
N_{90}	3.62	10.19	10.20
P_{90}	4.88	10.35	10.46
K_{90}	4.04	10.56	10.60
P_{90} K_{90}	4.85	10.32	12.24
N_{90} K_{90}	4.04	13.04	12.37
N_{90} P_{90}	4.68	12.76	12.24
N_{90} P_{90} K_{90} . . .	4.85	—	11.64
N_{60} P_{90} K_{90}	3.72	18.75	14.00
N_{120} P_{90} K_{90} . . .	4.42	16.20	15.60
N_{90} K_{90} P_{60}	4.12	10.44	14.37
N_{90} K_{90} P_{120} . . .	4.80	—	11.16
N_{90} P_{90} K_{60}	4.07	10.35	10.44
N_{90} P_{90} K_{120} . . .	5.78	9.48	11.68
Average . .	4.35%	11.89%	11.96%

mination of rubber content were taken before the setting in of frosts, on the 30th of October, 1942.

The yield of fresh roots at this moment was 21.5 centners per hectare of the experimental field; the rubber content was 4.91 per cent to the dry weight. The next sampling was done on the 12th December, 1942, after the soil was frozen to a depth of 25—30 cm; the roots were dug 18 cm deep. After the soil had thawn, the roots were placed into a thermostat and heated to 100° C during 30 minutes, after which they were dried at 50—60° C, and the rubber content determined. It should be noted that at the time when the samples were taken a marked secretion of latex could be observed. The rubber content at this time attained 6.67 per cent.

After ten days, on the 22th of December, 1942, the third sampling was done. Owing to a further sharp drop of temperature, the secretion of latex ceased, but the rubber content rised to 9.4 per cent. In the samples taken on the 12th and 26th of January, 1943, the rubber content was 8.57 and 8.90 per cent, respectively, *i. e.* only an insignificant decrease could be observed.

The biological processes taking place in the roots of kok-saghyz do not sease up to a certain moment even with freezing temperatures within the soil.

After the winter pause the study was not resumed until immediately before the beginning of the growing season, on the 10th of April, 1943. Since then samples were taken every ten days. These studies have revealed a number of facts which proved to be of interest, both theoretically and from the practical point of view. In the spring, at the beginning of the second season of life, a marked decrease in the rubber content takes place; probably because of its consumption for the formation of the plant organs. The rubber content increases again at the stage of the intense assimilation only, when it attains its maximum value—13.94 per cent to the dry weight. This condition is attained at the time of the full development of the rosette, appearance of the first flower buds and beginning of blooming. This occurred in our experiment on the 12th of May, 1943. At this time the roots of kok-saghyz contained the maximum amount of water, namely 87.44 per cent. Later, at the stages of flowering and fructification the rubber content showed a gradual decrease. At the close of fructification, on the 20th of July, the rubber content increased again, although never attained the original figure. Thus, for example, the roots of kok-saghyz contained 7.06 per cent of rubber on the 10th of April, 1943, at the time of

thawing of the soil. At the time of the appearance of the first vegetative organs, on the 21th April, the rubber content was 6.85 per cent; it attained 8.84 per cent at the time of the initial formation of the rosette, on the 2nd of May; on the 12th of May, at the time of the full development of the rosette, the rubber content was 13.94 per cent; at the time of flowering, on the 23th of May, it was 13.79 per cent; on the 3rd of June, the rubber content was 12.5 per cent; at the beginning of fructification, on the 14th of June, it was 9.18 per cent, and at the end of fructification on the 20th of July, 11.22 per cent. This is to say that the course of variation was the same as in the previous case.

On the 23rd of May, the yield was recorded; it was 23.1 centners of raw roots per hectare. The yield did not consequently decrease in spring. The death of roots and their decomposition within the soil which has often been referred to in the literature as being responsible for a decrease in yield and rubber, was observed by us only since the middle of June. This process reached its climax in July, so that towards the 20th of July the yield of roots had nearly decreased to half of the original. It is of interest that our experiment has shown that sharp changes in rubber content occur within short periods, *e. g.* within ten days from the 2nd of May to the 12th of May the rubber content increased from 8.84 to 13.94 per cent. This should be taken to mean that the period for harvesting kok-saghyz should be in each case chosen by means of laboratory analysis, for rubber content make widely change at different periods of the plant's life.

This conclusion is supported by other studies made by us on the accumulation of rubber in kok-saghyz grown under conditions differing from those which existed in our first experiment.

Conclusions. 1. In kok-saghyz rubber is present in a mobile form and evidently serves as reserve nutrient material which is easily consumed by the plant at the critical moments of its growth. This should in the first place account of the decrease in rubber content in the roots of kok-saghyz towards the beginning of the growing period and during fructification.

2. Kok-saghyz is highest in rubber by the second spring, at the time of the full development of the rosette. At this time, moreover, the greatest yield of rubber per unit area is obtained, for the death of the plants, which reduces the yield of rubber on the second year, occurs later, in the later half of June or in July. It should, therefore, be considered scarcely advisable to harvest the roots of kok-saghyz in the autumn of the first year.

3. Plantations which developed normally during the first season should be harvested in the spring of the second life's year, at the time when the rosette has attained its full development, which usually occurs in May. This will permit to use the field in the same summer for another crop, say for potatoes.

4. The harvesting of two-year-old roots of kok-saghyz at the end of July may be allowed for the purposes of seed production and in case the roots proved underdeveloped at the first year of plant's life.

5. The increase in the rubber content in the roots of kok-saghyz toward the beginning of winter should be explained by a partial conversion of the carbohydrates (in the first place of inuline) into rubber, owing to the action of the freezing temperature of the soil. Our studies have shown that a sharp drop in the carbohydrate content takes place at this time.

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Received
18. VI. 1944.

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Comptes Rendus (Doklady) de l'Académie des Sciences de l'URSS
1945. Volume XLVII, № 5

PLANT PHYSIOLOGY

DEVELOPMENT OF PLANTS AS AFFECTED BY NEGATIVE TEMPERATURES

By Z. A. KULCHITZKAYA

(Communicated by A. A. Richter, Member of the Academy, 8. VIII. 1944)

According to the theory of phasic development worked out by Lyssenko(1), along with other external factors, the plant requires a definite temperature to pass its first developmental stage. The effect of temperatures touching the limit of those necessary for the plant toward the end of the vernalization period is to modify its hereditary nature.

Studies by Abolina (2) and Eremenko (3) have shown the process of vernalization to be stimulated by the elements of mineral nutrition (N, P, K) and by irritants (ethylene, ether) administered at definite developmental stages.

The present study deals with the effect of low temperature acting upon the seeds of winter-crops subjected to various degrees of vernalization. Under experiment were winter-rye Omka, and winter-wheats *Lutescens* 0329 and Ukrainka. Seeds containing 37 per cent of water (to air-dry weight) and slightly sprouted were subjected to vernalization. In the one series of experiments the duration of vernalization was 2, 12, 22, 32 and 42 days; in the other series the duration was the same but, in addition, the seeds were chilled during 48 days after the period of vernalization. The latter was achieved by placing the container with the seeds into thawing snow, so that the temperature was kept constant at about 0°. The chilling temperature averaged −2° C, the extremes being from −0.5 to 4°C ($M = -2 \pm 0.03$). The seeds were planted in the open on the 22th of April, 1939.

The study of the primordial ear made on the 9th of June revealed important differences between the variants (Table 1).

According to the studies by Zabluda (5), Sapehin(5), Kruzhilin(6), Eremenko(7) and others, the differentiation of the primordial ear into spikelets is not possible until the vernalization stage is over. Observations by Saveliev (8) had shown that in the case of early dates of planting the primordial ear of the winter-wheat was fully differentiated in the autumn; with certain varieties it occurred even when planting was done late.

What is the degree of differentiation of the primordial ear which should be considered the initial one, still remains a matter of dispute.

In our experiment the setting of the primary primordial protuberances occurred in some variants prior to the close of the vernalization stage. As early as after two days of treatment the winter-wheat formed from 2 to 3 pairs of slightly expressed primordial protuberances; in rye they were 8 in number and clearly pronounced. In this experiment Ukrainka failed to form ears until autumn frosts set in, while in *Lutescens* 0329 only individual plants produced ears (Table 2).

Chilling of the seeds after two-day-long vernalization failed to produce any noticeable alteration in the development of the primordial ear in winter-wheat. The

Table 1

Degree of Development of Primordial Ears as Dependent on Different Duration of Vernalization and Chilling

Duration of treatment (days)	Winter-rye Omka		Winter-wheat *Lutescens* 0329		Winter-wheat Ukrainka	
	48 days chilling	no chilling	48 days chilling	no chilling	48 days chilling	no chilling
42	33.80 / 34	5.77 / 18	2.80 / 24	1.63 / 20	2.72 / 16	1.85 / 16
32	31.60 / 32	3.00 / 18	2.70 / 18	0.52 / 10	3.05 / 16	0.30 / 10
22	35.00 / 34	0.92 / 18	1.60 / 18	0.32 / 8	2.30 / 16	0.25 / 8
12	9.80 / 32	0.80 / 18	0.80 / 12	0.32 / 8	0.33 / 8	0.15 / 6
2	3.12 / 16	0.72 / 16	0.20 / 6	0.20 / 6	0.20 / 4	0.11 / 4

Note. The numerator is the length of the primordial ear (in mm); the denominator, the number of the primordial protuberances or spikelets.

Table 2

Time of Initial Earing (10 %) in Winter-Rye and Wheat as Related to Different Degree of Vernalization and Chilling

Duration of treatment (days)	Winter-rye Omka		Winter-wheat *Lutescens* 0329		Winter-wheat Ukrainka	
	48 days chilling	no chilling	48 days chilling	no chilling	48 days chilling	no chilling
42	15. VI / 47	20. VI / 53	2. VII / 63	3. VII / 66	28. VI / 61	1. VII / 64
32	15. VI / 47	22. VI / 57	30. VI / 63	12. VII / 75	27. VI / 59	9. VII / 72
22	19. VI / 51	27. VI / 59	2. VII / 65	26. VII / 89	27. VI / 59	25. VII / 88
12	24. VI / 56	28. VII / 90	10. VII / 72	27. VII single / 121	6. VII / 68	30. VII single / 124
2	26. VI / 58	3. IX / 127	26. VII / 88	3. IX single / 127	19. VII / 581	no earing

Note. Date of earing in the numerator; number of days till earing, in the denominator.

effect of chilling after 22, 32 and 42-day-long vernalization was considerably to accelerate the development of the plants, which told upon the growth of the primordial ear.

In rye the primordial ear was developing considerably faster than in winter-wheat. With 42-day-long vernalization followed by chilling the length of the developing ear in rye attained 33.8 mm, while in winter-wheats it was 2.72—2.80 mm only.

Under the conditions that prevailed in that year, 42-day-long vernalization proved fully sufficient for the completion of the vernalization stage, even in a variety possessing such a long first stage as *Lutescens* 0329. Single plants of Omka and *Lutescens* 0329, however, produced ears late in the autumn (on the 3rd of September) after two-day-long vernalization, *i. e.* thanks to the fact that they had passed ver-

nalization at a high temperature of the order of + 10°C, which prevailed at the beginning of spring and in autumn.

The effect of chilling applied after two-day-long vernalization was to enhance earing. In conformity with Lyssenko's experiments, this shows that at a negative temperature of the order of −2°C the process of vernalization does not cease, although its progress is but a slow one.

After a frost of −7°C (September 23), *i. e.* after a full arrest of growth, the plants which had formed ears were recorded (Table 3).

Table 3

Earing of Winter-Rye and Wheat in per cent on September 23

Duration of treatment (days)	Omka		*Lutescens* 0329		Ukrainka	
	48 days chilling	no chilling	48 days chilling	no chilling	48 days chilling	no chilling
42	100	100	100	100	100	100
32	100	95.1	100	69.2	100	42.8
22	100	74.2	100	20.0	92.1	27.8
12	83.6	18.9	69.0	3.1	76.4	4.7
2	91.6	11.8	16.8	1.0	9.6	0

Thirty-two-days of vernalization without chilling proved insufficient to bring about full earing. With the same time of vernalization followed by chilling, the plants came into ear at one, and ripening was attained at the earliest date. Thus for example, on the 12th of August there were 37.1 per cent of mature ears in *Lutescens* 0329 with 42-day-long vernalization, while the same period of vernalization followed by chilling resulted in 54.1 per cent of mature ears; with 32-day-long vernalization followed by chilling there were 77.5 per cent of ripe ears. Twenty-two days of vernalization with chilling resulted in a very protracted and irregular earing; in the variety Ukrainka it was incomplete.

Along with the above said, it could be observed from this experiment that the effect of chilling applied to seeds upon vernalization was to enhance the productivity of the plants when sowing was done in spring (Table 4).

Table 4

Yield of Winter-Rye and Wheat per 100 Plants in g (1939)

Duration of treatment (days)	Omka		*Lutescens* 0329		Ukrainka	
	48 days chilling	no chilling	48 days chilling	no chilling	48 days chilling	no chilling
42	150.9	144.1	107.0	103.0	110.5	93.5
32	157.2	92.7	112.0	74.0	114.5	76.1
22	105.7	31.4	88.0	33.0	127.6	20.3
12	85.1	0.7	59.0	0.9	—	—
2	78.7	2.7	8.3	0.4	3.2	0

With Omka and *Lutescens* 0329 the heaviest yield of corn was obtained in the case of 32-day-long vernalization followed by chilling, while in Ukrainka which has a shorter vernalization stage the same result was obtained in the case of 22-day-long vernalization followed by chilling. A comparison with the experiment involving 42-day-long vernalization without chilling showed the following increase in the yield of corn produced by chilling: in Omka, plus 13.1 g; in *Lutescens* 0329, plus 9.0 g; in Ukrainka, plus 34.1 g. In an analogous preliminary experiment car-

ried out in 1938 (the season was considerably colder and moister than in 1939) the increase in yield with 32-day vernalization followed by chilling was still higher than in 1939 (Table 5).

Table 5

Yield of Grain in Winter- and Summer-Wheats in 1938

Crop	Variety	Number of days		Yield of grains per 100 plants
		vernalization	chilling	
Winter-wheat	Ukrainka	42	13	193.2
» »	»	32	13	300.4
» »	*Lutescens* 0329	42	18	165.3
» »	» »	32	18	255.7
Summer-wheat	*Milturum* 0321	0	0	198.6
» »	*Caesium* 0111	0	0	194.6

The above data show that when sown in the spring winter-wheat surpassed the varieties (assigned to the given region) of summer-wheat after being vernalized during 32 days with subsequent chilling.

The effect of low (negative) temperatures applied along with vernalization should be considered as one of the conditions necessary for the development of plants. Negative temperatures of—2°C at a definite stage of vernalization enhance the development of winter crops and increase their productivity.

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Received 8.VIII.1944.

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Comptes Rendus (Doklady) de l'Académie des Sciences de l'URSS
1945. Volume XLVII, № 5

PLANT PHYSIOLOGY

CELL-DIVISION HORMONES IN BLOSSOMS OF COTTON-PLANT

By K. SUKHORUKOV and V. FILIPPOV

(Communicated by A. A. Richter, Member of the Academy, 9. VIII. 1945)

In our previous note (2) it was reported that vitamin B_1, as well as the yeast-hormone (bios), is synthesized by the cotton-plant in its leaves, accumulated in the stem, along with plastic substances, and spent during fruit-formation. It was also shown that the content of either of those substances in the stem may be changed by preventing photosynthesis and by removing blossoms or fruits. Analogous results were obtained by Bonner (3) for vitamin B_1 when girdling tomato stems. Our first communication (2) contained the results of experimental work carried on in 1939 at the Ferghana experiment station. The present note contains data obtained at the same station in 1940; this time we were kindly assisted in our experiments by the specialists of this station, D. I. Shmeleva and L. V. Rumshevich, to whom our thanks are due.

The blooming of the cotton-plant proceeds in a rather peculiar way, and the separate organs of the blossom display specific biological features (6). Our attention was mainly focussed on the study of the quantitative variation of the content of vitamin B_1 and yeast-hormone (bios) in the flower. But along with this, physiological observations were also carried out, bearing on the development and ageing of individual parts of the flower. In all cases as material use was made of *Gossypium hirsutum*, variety 8571, the same as had been investigated in our previous work (2) in July—August, 1940.

C o r o l l a. Upon the explosion of the flower, the development of the corolla in *Gossypium hirsutum* takes no more than a few hours, which makes this plant a convenient object for observing age variations. The table below contains figures relating to variation in dry weight of the corolla; for comparison figures showing the dry weight of ovaries are also given in this table.

Table 1

Variation in Dry Weight of Corolla and Ovary
(mean for 50 samples)

State of flower	Weight of corolla in g	Weight of ovary in g	Time of sampling
Buds, 24 hours before exploding	0.135	0.032	7 a. m.
Buds, 12 » » »	0.181	0.039	7 p. m.
Blossom at the moment of explosion . .	0.240	0.054	7 a. m.
» 12 hours upon explosion	0.188	0.065	7 p. m.
» 24 hours » »	0.155	0.605	7 a. m.
» 36 hours » »	0.130 (dies out)	0.080	7 p. m.

From the figures contained in Table 1 it may be seen that within the period of observation the weight of dry matter of the corolla exceeds that of the ovary considerably (1.6—4.6 times).

The run of variation fits within a single-peak curve, the maximum weight corresponding to the moment of the explosion of the flower.

The chemical alterations which take place in the corolla in the process of ageing are figured in Table 2. The characters compared are hygroscopicity, content of ether soluble substances («raw fat»), and total content of lipoids hydrolysed (according to Kumachava-Suto, without recalculation for glycerids). Corollae of different age were severed from the flowers, dried at 60°C, and upon drying and grinding, allowed to stay for 30 days in the laboratory, in glass-jars covered with thin layers of cotton. Dry weight was determined by desiccating at 105°C.

Ether extract from powder dried at 100°C was obtained with the aid of Soxlet's apparatus.

Table 2

Chemical Alterations in Corolla

State of flower	In per cent to dry matter		
	Hygroscopicity	Ether-soluble substances	Saponifiable lipoids (after Kumachava-Suto)
6 hours previous to explosion	9.40	4.48	—
12 » upon explosion	8.56	4.57	4.33
24 » » »	—	7.51	4.12
30 » » »	5.35	8.63	—

Neither plant physiology, nor phytochemistry gives us any clear idea as to the processes that go on and the condition that prevail in cells of different age ([1],[4]).

The decrease in hygroscopicity seems rather to be in agreement with Ružička's ([5]) interpretation of the hysteresis of plasmic colloids, which accompanies the ageing of organisms, or any treatment which accelerates it. A well-pronounced increase in the content of ether-soluble substances of free lipoids along with a constant total content of saponifiable lipoids testify to a process approaching that of lipophanerosis, if not identical with it. Lipophanerosis, it is true, has as yet been observed in rapidly decaying cells only, if affected with various poisons or other agents lethal to the cell. Theoretically, however, it is not altogether out of the question in the case of natural ageing or death either ([4]). The yeast hormone was determined according to the method previously described, but extraction was carried out this time with water, over a boiling water-bath. The results of the determinations are summarized in Table 3.

Table 3

Bios-Content in Corollae

State of corolla	Bios, in relative figures	Note
6 hours before explosion	12.62	Since the moment of explosion corollae severed from blossoms were kept in humid chambers; for analysis they were fixed by drying after 12 hours (young samples) and after 30 hours (old ones)
12 hours upon »	9.88	
30 » » »	6.89	
Severed corollae, young	9.36	
Severed corollae, old	9.60	

From the above figures it may be seen that the amount of yeast-hormone is decreasing rapidly, according as the corollae are ageing together with the flowers. Unlike this, in such corollae as were severed from the flowers, the content of the yeast-

hormone is found to remain constant. A comparison of the respective figures makes us conclude that the bios contained in the corollae is not spent by them in the process of their development, but migrates to other parts of the blossom or of the plant.

O v a r i e s. Ovaries of different age served us for determining the content of bios and vitamin B_1. Along with the determination of the content of hormones we determined also the content of protein nitrogen and the run of dry mass accumulation.

Each sample comprised 50 prepared blossoms; the severed ovaries were dried rapidly. Table 4 represents the results of these determinations, as recalculated to one ovary.

Table 4

Dry Weight Protein Nitrogen, Bios, Vitamin B_1 (per one ovary)

Time of sampling	Dry weight in g	Protein nitrogen in mg	Bios in relative units	Vitamin B_1 in relative units
24 hours before fertilization	0.030	—	9.60(100%)	11.27 (100%)
At the period of »	0.045	2.18(4.84%)	10.00 (70%)	12.17 (70%)
36 hours after »	0.081	3.80(4.69%)	14.40 (56%)	13.27 (43%)
84 » » »	0.163	6.67(4.10%)	18.27 (35%)	16.23 (27%)
240 » » »	1.482	38.16(2.57%)	44.80 (9%)	—)
360 » » »	2.917	75.54(2.56%)	57.33 (6%)	30.60 (3%)

From these figures it may be clearly seen that, according as the ovaries develop and grow larger; they accumulate all the substances here recorded; the relative contents of proteins, bios and vitamin B_1 are obviously decreasing, however. To make the results of our analysis given in Table 4 more illustrative, the same figures are given in Table 5 in relative units, the indices for the period of fertilization being put at unity (Table 5).

Table 5

Dry Weight, Protein Nitrogen, Bios, Vitamin B_1 (per one ovary)

Time of sampling	Dry weight	Protein nitrogen	Bios	Vitamin B_1
24 hours before fertilization	0.66	—	0.96	0.93
At the period of »	1.00	1.00	1.00	1.00
36 hours after »	1.80	1.74	1.44	1.09
84 » « »	4.62	3.06	1.83	1.33
240 » « »	32.93	17.50	4.48	—
360 » »	64.82	34.65	5.73	2.51

The figures of Table 5 bear out with clarity that the accumulation of the substances analysed proceeds in no regular way; thus, the non-proteinic substances are accumulated at the highest rate, whereas bios is stored much slower, and vitamin B_1 slower still. Attention is attached to the figures expressing the content of bios recorded soon after fertilization; it is at this time that the ovaries are highest in bios. Here a certain analogy may be traced between the cells of an animal and a vegetable organism.

It has been firmly established by Koegl and his co-workers that at the earliest stages of their development animal embryos are very rich in biotine. Similar results have been obtained by us for vegetable organisms. It seems therefore that bios is playing an important rôle in the embryonic cells of both animal and vegetable organisms, though the precise nature of this rôle is still obscure to us.

Conclusions. Corollae of the cotton-plant are a convenient object for physiological and chemical investigation of age variation in the cell. Along with ageing, a translocation of dry substances and bios is observed from the corolla to other parts of the flower and of the plant.

The accumulation of total dry mass, proteins, bios and vitamin B_1 was observed to take place throughout the development of ovaries. Young ovaries have been found to have the highest content of bios relatively to the total mass. A certain similarity in the processes of bios accumulation in animal and vegetable embryos has been observed.

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Received
3. VIII. 1944.

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Comptes Rendus (Doklady) de l'Académie des Sciences de l'URSS
1945. Volume XLVII, № 5

MICROBIOLOGY

THERMOPHILIC VARIANTS OF BUTYRIC-ACID BACTERIA

By A. IMŠENECKI and L. SOLNTZEVA

(Communicated by V. L. Komarov, Member of the Academy, 3. I. 1945)

The decomposition of organic matter by microorganisms is enhanced according as the temperature of the medium rises. This increase in the rate of the process of mineralization depends on the activity of the thermophilic microorganisms and is limited by their temperature maximum of growth. For a successful disjunction of such bioceonoses as are active at high temperatures it is necessary to work out a technique for the isolation and selection of the most active forms; however, no subsequent utilization of these pure bacterial cultures is possible unless their physiology is sufficiently known.

In the course of our researches on the biology of thermophilic bacteria we turned to the study of anaerobic carbohydrate fermenting bacteria. In searching for such forms we found that they are but rarely encountered under natural conditions, belonging in their majority only to heat-tolerating forms rather than to true thermophiles. Their temperature optimum of growth, though they are able to develop at 55°C, is considerably below this figure. Transferred to high temperatures, these strains degenerate and sometimes die out. It was necessary to try to find such conditions under which mineralization of organic substance is going on continuously at high temperatures and, consequently, the selection of various thermophiles has already taken place.

There is one field of utilization of the activity of thermophilic bacteria in which such conditions are available. This is the fermentation of the sediment of sewage in the so-called thermophile methane-tanks. The most variegated composition of this sewage sediment, including proteins, carbohydrates, fats and other substances, made it promising for our purpose. Considering that in the methane-tanks conditions of reduction are prevalent, strict and facultative anaerobes alone were isolated. At first the sediment was seeded only into a liquid medium of the following composition: potatoes (cut into small bits), 20 g; chalk, 2 g; tap-water, 100 g. However, this medium was found unsuitable for the development of anaerobic bacteria: fermentation was slack and in subsequent transfers stopped altogether. Taking into account the usual composition of the medium of the thermophiles, filtered sewage (50 per cent) and yeast-water (50 per cent) were substitute in it for tap-water. In this substrat the bacteria gave a profuse growth.

Various anaerobic bacteria were thus isolated from a methane-tank working at a high temperature. Among these were three strains of one and the same species which readily fermented potatoes with the production of gas, and were strict anaerobes. The bacteria of this species may be described as follows.

Their vegetative cells are the shape of rods with rounded ends, their size is: 1.0 $\mu \times 3.5$—7.4 μ. In young (3—5-hour-old) cultures the cells are actively mobile, but in 17-hour-old cultures this mobility is already lost. Quite soon the vegetative

cells begin to grow considerably thicker and assume the shape of a spindle. Granulose, staining blue after iodine treatment, appears in them. These changes correspond to the beginning of sporulation, and at this stage of their life-cycle the bacteria present a very typical aspect: an unstained prospore is visible at one end of the spindle, while the rest of the cell is stained blue.

The ripe spore is oval in shape and measures 1.3×1.7 μ; it occupies in the cell a subterminal position. If the spores are transferred into a fresh medium, they germinate polarly, the entire process of germination being completed within a relatively short time (1.5—2 hours).

The bacterium is an anaerobe, and accordingly it develops on the surface of solid media in a vacuum-desiccator only. On an agar medium, composed of glucose (2 g), sewage filtrate (100 g), agar (2 g), it produces round, smooth, gossy; transparent colonies of a greyish-yellow tint. The colonies are convex with an even margin and reach 0.5—0.8 mm in diameter. In potato-glucose agar in test-tubes (in high «columns») lens-shaped colonies are formed, and the agar is ruptured by the gases produced.

Individual colonies treated with iodine stain black-blue, the cells composing the colony containing granulose. Although the species described is anaerobic and cannot grow on the surface of solid media under ordinary conditions, it is nevertheless not so pretentious as to conditions as many other anaerobic species of bacteria. This may be seen from its ability to develop, when grown in a liquid medium, in a relatively shallow layer of liquid, some 4—5 cm deep.

The described species does not grow under anaerobic conditions on such commonly used media as meat-peptone agar or broth, mash, mash-agar, potato-agar and sliced potatoes. A good growth is obtained in a medium consisting of potato broth, gelatine and sewage filtrate, but gelatine is not liquefied.

As has already been pointed out, the bacterium does not develop readily unless in media added with filtered sewage. Sewage filtrate to which various alcohols and carbohydrates in a quantity of 2 per cent were added, was, therefore, tested for its ability to ferment these substances. The bacterium was found to ferment glucose, levulose, galactose, dextrin and starch with acid and gas production, the gases produced being CO_2 and H_2. A number of other substances, such as sucrose, maltose, lactose, raffinose, arabinose, mannitol, glycerine, inulin, pectin and cellulose were not fermented by the bacterium.

Tests with fermentation of 20 per cent potato-mash were arranged with the purpose of determining the nature of the products forming in the process of fermentation. The fermentation was carried on in 500 cm^3 flasks for 48 hours at 55°C. Among the products of fermentation butyric and formic acids were found, as well as ethyl alcohol.

The species isolated does not grow at all at 27°C, *i. e.* it is a true thermophile. At 37°C its growth is delayed, while most intensive growth is observed at 53—55°C. The spores of the bacterium are relatively heat-resistant and survive boiling for 30 minutes, but are killed within the same period of time at 112°C.

The anaerobic bacterium described above may be classed with the butyric-acid group of bacteria. The latter are rather inadequately united in one species designated as *Clostridium butyricum* or *Bacillus amylobacter*, quite different species, such as, *e. g. Granulobacter pectinovorum*, *Clostridium acetobyturicum*, *et al.* being residually included into it. The described bacterium is closest to *Clostridium Pasteurianum* first described by Winogradsky, and we are inclined to regard it as an earlier unknown thermophilic variant of this species. The particular character common to both variants is the spindle shape of their cells stained blue by iodine (test for granulose) during spore-formation.

From the «basic» mesophile form of *Clostridium Pasteurianum* our variant is distinguished not alone by its high temperature optimum, but also by its high temperature minimum of growth. The mesophilic forms may develop even at 10°C, while the thermophilic variant does not grow below 27°C. Accordingly, the spores of the thermophile are more resistant and survive boiling during 30 minutes, while the spores of the mesophilic variant are killed in 10 minutes at 100°C.

The physiology of nutrition of the thermophilic variant differs considerably from that of the common form, since the development of the thermophile and the fermentation of carbohydrates by it is possible in media containing sewage filtrate, *i. e.* the same medium from which it has been isolated. It seems to have adapted itself to life in a medium very rich in proteins and other organic substances and, quite possibly, in supplementary factors as well. This fact represents a useful example of the ecological physiology of microorganisms.

Thus, this thermophilic race of butyric-acid bacteria takes part, alongside of other anaerobic forms, in the mineralization of organic matter in methane-tanks operating at high temperatures. The intensification of this process observed at relatively high temperatures depends naturally on the high biochemical activity of thermophilic bacteria and of the ferments excreted by them into the surrounding medium.

From the theoretical viewpoint it is of interest that, owing to the plasticity of microbes, dissociation and selection give rise to thermophilic variants in definite systematic species. Analogous «parallel» forms are observed in anaerobic cellulose decomposing bacteria as well, and the possibility cannot be excluded the existence of thermophilic variants arising by way of adaptation may yet be observed in most different species of microorganisms. The microbes may become adapted to various conditions offered by the surrounding medium. Suffice it to remind of the osmophilic, acidophilic and thermophilic microbes. However, it is hardly right to establish, as this is continuously done, special thermophilic species of bacteria, without having found out beforehand if there are no species taxonomically analogous to them.

In those cases where it is possible to establish the simultaneous existence of both mesophilic and thermophilic variants, one should think of the secondary origin of the thermophiles. In this connexion the reproduction of such an adaptation seems to be of interest, especially in the case where the forms needed cannot be isolated from natural substrata.

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Received
3. I. 1945.

Comptes Rendus (Doklady) de l'Académie des Sciences de l'URSS
1945. Volume XLVII, № 5

MICROBIOLOGY

A NEW BACTERIUM ON THE RUBBER-PLANT TAU-SAGHYZ

By A. N. KLECHETOV

(Communicated by A. A. Richter, Member of the Academy, 13. VII. 1944)

The results of our experiments of 1941 (1) made us think of the possible ability of tau-saghyz to assimilate gaseous nitrogen. The experiments of 1942 have also shown that this rubber-plant is able to develop in sand and water cultures on composite nutrient substrata lacking nitrogen (see table).

Variants of sand cultures	Dry weight per container		
	of above-ground part, g	of roots, g	relative values
Tau-saghyz on composite substratum without nitrogen	0.14	0.38	100
Tau-saghyz on complete mineral composite substratum	0.26	0.67	185
Peas on composite substratum without nitrogen	6.4	0.82	100
Peas on complete mineral composite substratum	36.45	1.4	569
Oats on composite substratum without nitrogen	5.71	2.2	100
Oats on complete mineral composite substratum	53.11	2.66	930

The least decrease of yield on substrata without nitrogen was obtained, like in our earlier experiments, with tau-saghyz (1.8 times in comparison with the control plants on complete mineral substratum). The yield of peas without nitrogen was decreased 5.7 times; that of oats, 9.3 times.

The tau-gaghyz plant itself has a high nitrogen content. Various specimens of the grain of this rubber-plant contained from 3.49 to 3.86 per cent of total nitrogen. The roots of tau-saghyz from different variants of pot-culture experiments contained from 2.19 and 3.96 per cent of nitrogen; its aboveground mass, between 2.32 and 4.52 per cent.

In examining microtome sections of the mycorrhyze my attention was drawn by a spherical bacterium encountered by me in preparations of perfectly sound plants (from the plantation or from pot-culture experiments), alongside of the fungus. Direct microscopic examination shows that this microorganism is present in every investigated organ of tau-saghyz (in its roots, peduncle, leaves, heads and seeds) and is particularly abundant in the tips of young roots. Similar findings were made by me as early as in 1937—38, but at that period the pure cultures of the microbe were not obtained.

The essential method of the work of Henning and Villforth [2] consisted in isolating the bacteria from the internal tissues of various organs of a number of plants. But these authors do not desribe the microbes isolated by them in detail.

On June 25, 1943, I isolated a pure culture of the bacterium found by me in tau-saghyz from sound tips of young seedlings.

The micrococcus described grows readily on nitrogenless mannitol agar prepared on distilled or tap-water. The composition of the medium was as follows: water, 100 cm^3; agar, 2.5 g; mannitol, 2 g; K_2HPO_4, 0.02 g.

Rhodococcus corallinus, Levine et Soppeland, *Proteus vulgaris* Hauser and a species of wild yeasts seeded in the same medium, with the purpose of comparison, did not grow in it.

The following characters of the new microbe have been established.

Its cells are spherical, 0.4—0.8 μ in size; they lie singly or in pairs, or in short chains. Its colonies on agar made with broth prepared from tau-saghyz seedlings are small, rounded, droplet-like, of a greyish-white tint, with an even margin (24-hour growth); their surface is smooth; the colonies are semi-pellucid.

Slanted agar (tau-saghyz broth or chickory root broth): continuous growth along the streak at the bottom of the test-tube, fine granulary growth higher up, the margin somewhat uneven (48-hour growth); mucous growth flowing downwards.

Stab-culture in tau-saghyz agar: ready growth on the surface, somewhat weaker in the upper part of the stab, scarcely any growth in its lower part.

Mannitol agar without nitrogen: ready growth, but slower than in tau-saghyz broth agar; the character of the colonies is similar; no acid production.

Meat-peptone agar: a very fine film after 24 hours, a fine semi-pellucid film after 72 hours; at the bottom of the test-tube continuous, in the upper part of the slant a fine granulary growth; greenish fluorescence.

Potatoes: yellowish film after 24 hours of growth; after 48 hours, a yellowish film; the potato-slice becomes dark along the string, offensive odour appears; after 72 hours a yellowish raised growth, the slices have become dark; after 7 days, the streak spreads.

This microbe is possibly able to fix atmospheric nitrogen, but additional study is necessary to prove this. It is aerobic. It grows readily at 25°C. Habitat: roots and other organs of the rubber-plant tau-saghyz.

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Received
13.VII.1944.

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Comptes Rendus (Doklady) de l'Académie des Sciences de l'URSS
1945. Volume XLVII, № 5

MICROBIOLOGY

THE AETIOLOGICAL AGENT OF A PECULIAR FORM OF ENCEPHALITIS

By N. I. KHODUKIN and M. N. SOSHNIKOVA

(Communicated by E. N. Pavlovsky, Member of the Academy, 20. VII. 1944)

In the winter of 1942—43 an epidemic of an unknown disease, characterized by high lethality, broke out in one of the districts of the Tashkent region. It was established by E. A. Galperin that alongside of other diseases, cases of encephalitis were prevalent in this outbreak. Dr. Sirota also came to a similar conclusion on the ground of path-anatomic material.

People's Commissariat of Health sent accordingly an expedition with the purpose to determine the aetiology of this encephalitis. The virusological group of the expedition worked with intervals from January to June of 1943. The supposition was advanced that this encephalitis was of toxic origin. This hypothesis was rejected, however, after a careful examination of the food which had been used by the victims. In order to clear up whether or not the encephalitis cases might be due to typhus, the brain of persons who had died from the disease and the blood of patients was inoculated to 14 guinea-pigs and 30 rabbits; none of the guinea-pigs showed the typical typhus temperature curves; neither did any of the rabbits which received 5 cm^3 of brain emulsion of persons who had died of encephalitis, or of blood or liquor from encephalitis patients, gave a positive Weil-Felix test. No agglutinin rise was detected in the course of the systematic Weil-Felix tests of the sera of patients.

The third hypothesis was that this encephalitis was caused by a filtrable virus. For the purpose of testing the virusaetiology of the disease, the following plan of investigation was adopted. White mice were used for the experiments; they were infected intracerebrally (by the introduction of 0.02 cm^3 of infective material) according to the usual technique: a) with the blood of patients obtained during the first few days of illness; b) with the liquor of same; c) with the brain of persons who had died of encephalitis; d) intranasally with washings from the nasopharynx of patients, the brain of the intranasally infected mice being subsequently used for infecting fresh mice intracerebrally. After two subsequent passages the results were considered negative.

Simultaneously the patients in every case were subjected to a bacteriological investigation by V. I. Pissarenko. Blood was inoculated into neutral and alkali sugar broth and into liver broth under anaerobic conditions. The brain was investigated in a similar manner. No bacteriological findings were in any of the typical cases of encephalitis. Patients with a negative Well-Felix reaction were selected for the infection of mice. The same condition was observed in the selection of material from corpses. Although the exact day of the beginning of the illness could not be always strictly established because of the peculiarities of the mode of life of the local population, we strived to obtain the infective material at the earliest date possible. Blood was obtained on the 2nd, 5th, 7th and 9th day of the disease, liquor on the 10th day.

In some cases the encephalitis took a very rapid course and led to a fatal end during the first days of the disease, therefore we were able on several occasions to inoculate mice with the brain of persons who had died a few days after the onset of the disease. In the course of this work a total of 24 cases were investigated: 18 patients and 6 corpses. In a part of cases, in which the mice did not show evident symptoms of disease, the passages were discontinued. These cases are not included into the table, in which the results are summed up (Table 1).

Table 1

Material used for infection of mice	Day from onset of disease	Number of cases	Number of strains of virus isolated			
			on first inoculation	in first passage	in 2nd passage	total
Blood	2, 5, 7, 9	5	2	1	2	5
Brain	2, 3, 5, 7, 8, 12	7	1	1	2	4
Liquor	2, 5, 10	4	1	1	2	4
Nasopharyngeal washings	2, 5	2	—	1	1	2

As a rule, the mice fell ill on the 4—5th day after intracerebral inoculation. They became limp, their hair standing on end. Quite often erethism of the animals could be observed. A few hours before death paresis and paralysis, mostly of the posterior limbs, set in. In many cases general convulsions developed at the approach of death, and the mice usually died during this convulsive fit. Often already the primarily inoculated mice fell ill and died in high numbers. In the first passage over 40 per cent of the intracerebrally inoculated mice died. The percent of dying mice increased with every new passage from mouse to mouse, and by the 5th passage the death rate reached 100 per cent. At a later stage of our work two strains were selected for a detailed study.

The first strain was isolated from the defibrinated blood of patient R., with which the mice had been inoculated intracerebrally on the 9th day of her disease. 4 days later 4 mices fell ill. A first passage was made from one of these on 6 fresh mice of which 5 died by the 7th day. Already in the second passage all the infected mice died. By the present moment the strain is in its 33rd passage. From the 24th passage onwards a 100 per cent mortality from a dose of 0.03—0.02 cm^3 of a 10 per cent suspension of infected mouce-brain became established.

The second strain was obtained by inoculating mice with the suspension of the brain of patient H. who died on the 7th day from the onset of the disease. Bits of brain were taken from *corpus striatum* and the internal capsule, and emulsified in neutral broth. This emulsion was introduced intracerebrally to 5 mice. By the 4th day 2 of the mice were dead. In the first passage already 7 mice out of the ten used for it died showing the characteristic symptoms of encephalitis: paresis, paralysis and generalized convulsions. By the 16th passage the strain killed 100 per cent of the mice, when inoculated into the brain in a dose of 0.02—0.03 cm^3 of brain emulsion. In the course of this passage the virus was contaminated with some cocci. In order to remove them it was decided to filter through a Seitz filter (EK) the brain emulsion of mice that had died or had been killed at the agonal stage. Before filtration the emulsion was diluted 100 times. After the sterility of this filtrate was ascertained by bacteriological tests, it was used for the inoculation of mice. This experiment showed that the virus readily passes through the Seitz filter without any noticeable decrease of virulence. By the present moment the 2nd strain is already in its 31st passage.

Having ascertained that both strains produce identical results as regards the infection of mice, the controls showing perfect bacteriological sterility, we turned to neutralization tests with the sera of convalescent patients. The technique used in these tests was the one usually employed. The results of the first neutralization tests gave little hope of success. They showed that virulicidal substances do not ap-

pear in the sera of convalescents in sufficient quantity, at any rate until the 70 days after the onset of the disease. The next series of tests was made, therefore, with the sera of convalescents, beginning approximately from this date of their disease. The technique in these tests was similar to that of the first. Their results are summed up in Table 2.

Table 2

Serum	Virus dilution					
	1:100	1:1000	1:5000	1:10000	1:25000	
1. J. (140 days from onset of disease)	3/1	3/1	3/1	3/0	—	Virus strain 1
2. Faz. (over 100 days from onset) . .	3/2	3/0	3/2	3/0	—	Virus strain 1
3. P. (105 days from onset)	3/0	3/0	3/0	3/0	3/0	Virus strain 1
4. R. (95 days from onset)	3/0	3/0	3/0	3/0	3/0	Virus strain 1
5. Faz (cf. 2)	3/2	3/1	3/1	3/0	3/0	Virus strain 2
6. J. (cf. 1)	3/1	3/1	3/1	3/1	3/0	Virus strain 2
7. Control (normal serum)	3/3	3/3	3/3	3/3	3/3	Virus strain 2

Note. Numerator, number of inoculated mice; denominator, number of deaths.

The results of the above neutralization tests show with certainty that we had to do with a virus pathogenic for man. Thus, the virus nature of the epidemic of encephalitis studied by us may be considered proved. By the present time it has moreover been established that the virus does not lose its pathogenicity for mice for at least 2 months, if it is stored in glycerol in the ice-box (at + 4°C). Observations concerning the time of the circulation of the virus in the human organism show that up to 3 months from the beginning of the disease it may still be isolated from nasopharyngeal washings. Rabbits and guinea-pigs seem to be unsusceptible to this virus.

Its position among the other viruses which provoke encephalitis remains yet to be investigated. The sera of verno-aestival encephalitis do not neutralize it. The epidemiological characteristics of the epidemic in the course of which the virus described in the present report has been isolated, induces us also to presume its being different from any other known virus of encephalitis, at least from those encountered within the limits of the USSR. Its particular characteristics may be resumed as follows. Cases of the disease appeared in winter. Maximum prevalence was reached in March—April. In May—June the epidemic was extinct. Fresh cases appeared in the autumn of 1943, but their number decreased abruptly falling to 2 cases in December.

Clinically this encephalitis differs considerably from both types prevalent in the Far East, the aestivo-autumnal and the tick-borne verno-aestival ones (L. J. Shargorodsky and J. J. Gordon). This encephalitis differed also from the Alma-Ata type (a tick-borne encephalitis). The contagiousness of our encephalitis is very low. The narrowness of the territorial limits within which the disease is prevalent is one of its prominent characters. This last circumstance induces to seek a natural reservoir of the virus. The possibility that it is transmitted by ticks is by no means excluded. In this connexion the ticks of the genus *Ornithodorus* seem to arouse the greatest suspicions. In the zone of the prevalence of encephalitis they occur in great numbers. Ticks of the species *Ornithodorus papillipes* which had been fed on an encephalitis patient, transmitted by their bites the disease to white mice. The virus transmitted thus through tick-bites was proved by means of neutralization tests to be identical with the virus isolated directly from human cases.

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Received
20.VII.1944.

Comptes Rendus (Doklady) de l'Académie des Sciences de l'URSS
1945. Volume XLVII, № 5

HELMINTHOLOGY

ON THE MORPHOLOGICAL NATURE AND TAXONOMIC VALUE OF NEMATODES BELONGING TO THE GENERA *DICHEILONEMA* DIES., 1861, AND *MONOPETALONEMA* DIES., 1861

By K. I. SKRJABIN, Member of the Academy, and N.P. SHIKHOBALOVA

During the two last decades in the helminthological literature devoted to the morphology and taxonomy of nematodes of the family *Filariidae*, two genera were often dealt with, *viz.* the *Dicheilonema* and the *Monopetalonema*, both founded by Diesing as far back as in 1861 (4). The typical characters of these genera are interpreted by the various authors in a different way; as a result, the criterion necessary for a distinct determination of their taxonomic position has nearly been lost.

In the course of a monographical treatment of the family *Filariidae* we not only become aware of, but even were confronted with those conflicting data which abound in the literature on *Dicheilonema* and *Monopetalonema*. And ultimately we found ourselves able to shed revealing light upon the morphological nature and the taxonomic value of these representatives of the family *Filariidae*.

Monopetalonema Diesing, 1861

In 1861 Diesing(4) described the genus *Monopetalonema* for two nematode species. *Filaria physalurum* Bremser, 1851, from *Alcedo amazona*, and *Filaria obtuso-caudatum* Dies., 1861, from *Laniidae*. The latter species is notably identical with *Filaria nodulosa* Rud., 1820.

During the several decades that followed Diesing's death the genus *Monopetalonema* was no subject of attention on the part of helminthologists. Consequently, in the review of Yorke and Maplestone (5) this genus is mentioned under the heading of «*Filariidae* insufficiently known».

These authors give the following diagnosis of the genus *Monopetalonema*, which decidedly lacks any precision: «*Filariidae*, body very long, filiform; mouth with two labiae; male: caudal end bent ventrally; caudal alae and papillae present; spicule single! Female: blunt straight tail vulva in the anterior part of the body (?). Parasitic of connective tissue and abdominal cavity of Brazilian birds and badgers. Type-species:?*M. physalurum* (Bremser, 1851) Dies., 1861. The second species, *M. eremita* Leidy, 1886, from *Meles labradorica*».

The diagnosis suggested by Yorke and Maplestone is accompanied with the following note: «The type species of this genus is uncertain. Diesing placed in it two species, *viz. M. physalurum* (Bremser, 1851) and *M. obtuso-caudatum* Diesing, 1861; the latter is synonymous with *Filaria nodulosa* Rud., 1820, and belongs to the genus *Contortospiculum*. Very little appears to be known regarding *M. physalurum* (Bremser, 1851), but it seems quite possible that it really belongs to the same genus as *M. obtuso-caudatum*, in which case the genus *Contortospiculum* would fall a synonym of *Monopetalonema*».

In spite of this remark, Yorke and Maplestone do not object against the genus *Contortospiculum* Skrjabin, 1915, as an independent genus, in which they include *Filaria nodulosa* Rud., 1820 (syn. *Filaria obtuso-caudata* Dies., 1861).

Such was the state of affairs till 1926. In 1928 appeared a paper by Skrjabin and Petrov(7) who, upon having thoroughly investigated *Filaria nodulosa* Rud., 1820, arrived at a conclusion that this species should be excluded from the genus *Contortospiculum* and separated as the typical species of a new genus to which they gave the name of *Parhamatospiculum*.

Thus, one of the components of the obscure genus *Monopetalonema*, namely, *Filaria nodulosa*, received a definite place in the system of *Filariidae* as a representative of the subfamily *Setariinae*.

In the same year (1928) appeared Boulenger's paper(3), containing a revision of the study of *Monopetalonema physalurum* (Bremser, 1851) which had previously been considered as the type-species of the genus *Monopetalonema* Diesing, 1861. Boulenger's conclusion was that *Monopetalonema physalurum* is undoubtedly a representative of the genus *Politospiculum* Skrjabin, 1916, which should in its turn be taken to mean that the genus *Politospiculum* is synonymous with *Monopetalonema*.

In this way the analysis of the two species that had been united by Diesing into a single genus, *viz. Monopetalonema*, showed that this genus had injudiciously been made to include not only representatives of two different genera, but even of different subfamilies, in so far as *Filaria nodulosa* is the representative of *Setariinae*, whereas *F. physalurum* belongs to the *Filariidae*.

We are thus ultimately led to the following state of affairs: is *Filaria nodulosa* be considered as the type-species of the genus *Monopetalonema*, then the genus *Parhamatospiculum* should be regarded as synonymous with *Monopetalonema*; if, on the other hand, we consider as the type-species of the latter genus *Filaria physalurum*, it is the genus *Politospiculum* that has to be regarded as the synonym of *Monopetalonema*.

In order to avoid further confusion resulting from the mistake made by Diesing who had united in the genus *Monopetalonema* two species belonging to different subfamilies of *Filariidae*, we deem it expedient to introduce the following corrections into the taxonomy of *Filariidae*:

1. The genus *Monopetalonema* Diesing, 1861, as being no consistent taxonomic unit, is to be excluded from the taxonomy of *Filariidae*.

2. One of the components of the genus *Monopetalonema*, the species *Filaria nodulosa* Rud., 1820, has to be referred to the genus *Parhamatospiculum* Skrjabin et Petrov, 1928 (subfamily *Setariinae*).

3. The second component of the genus *Monopetalonema*, the species *Filaria physalurum* Bremser, 1851, has to be referred to the genus *Politospiculum* Skrjabin, 1916 (subfamily *Filariinae*).

4. The question as to the generic position of *Filaria eremita* Leidy, 1886, from *Meles labradorica*, which by Yorke et Maplestone (1926) was referred to the genus *Monopetalonema*, is to be subjected to closer examination.

Genus *Dicheilonema* Diesing, 1861

This genus was founded by Diesing and made by him to embrace ten nematode species (4).

In 1915 Skrjabin(1) founded the genus *Contortospiculum* with a type-species *C. rheae* (Owen, 1843), a nematode from the serous cavities of *Rhea americana* and the *Otidae*. The second species referred by him to this genus is a nematode which was described in the end of XVIII from the stork as *Filaria ciconiae* Schrank, 1788 (synonymous with *Filaria labiata* Creplin, 1825).

According to the opinion expressed by Stiles and Hassall(8), *Filaria labiata* Creplin, 1825, should be regarded as the type-species of the genus *Dicheilonema*

Diesing, 1861. If this view be adopted, the genus *Contortospiculum* will thus be reduced to the rank of a synonym of *Dicheilonema*. As a matter of fact, Railliet ([6]) writes that Diesing has founded the genus known by his name solely on account of the tubular shape of the large spicule, a character typical of *Filaria bifida* Molin, 1858, alone, which has also been referred by Diesing to the genus *Dicheilonema*. All these considerations made Railliet arrive at the conclusion that it is *Filaria bifida* Molin, 1858, that ought to be regarded as the type-species of the genus *Dicheilonema*.

If we are to adopt this point of view, the genus *Contortospiculum* cannot be considered synonymous with *Dicheilonema*. This opinion is shared by Yorke and Maplestone ([9]), according to whom the following three species have to be referred to the genus *Contortospiculum*: *C. ciconiae* (Schrank, 1788), *C. rheae* (Owen, 1843) and *C. americanum* Railliet, 1918, from *Asio accipitrinus*. These English authors place the genus *Dicheilonema* among the category of «*Filariidae* insufficiently known», with *Filaria bifida* Molin, 1858, from *Dactylomys amblyonyx* as its type-species. Baylis and Daubney (1926) adhere to the opinion expressed by Stiles and Hassall, considering the genus *Contortospiculum* as a synonym of *Dicheilonema*. The same point of view has been advocated by Walton (1927) and Boulenger ([3]).

In 1939 Freitas and Lent ([5]) described a new *Filariidae* genus, *Molinema* n. g., on the basis of three nematode species, two of which had previously been described by Molin, *viz.*: *Filaria diacantha* and *F. bifida*. Thus, an attempt was made to include into the genus *Molinema* a parasite which Railliet (1918) considered to be the type-species of the genus *Dicheilonema*.

It stands to reason that such genera as *Molinema* (with the species *M. bifida*) and *Contortospiculum* (with the species *C. ciconiae*), which are both well-differentiated and quite independent of each other, leave any identification out of the question. From what has been said above we feel authorized to draw the following conclusions:

1. The genus *Dicheilonema* comprises obviously heterogeneous species which have to be referred to different independent genera.

2. The question as to the type-species of the genus *Dicheilonema* is answered in a different way by different authors: Stiles and Hassall([8]) suggest as typical species *Filaria labiata* Creplin, 1825, whereas Yorke and Maplestone argue in favour of *Filaria bifida* Molin, 1858.

3. It is therefore impossible to restore the precise diagnosis of the genus *Dicheilonema* in the form it had primarily been conceived by Diesing ([4]).

4. Hence the necessity to pronounce *Dicheilonema* a genus of mixed character, which has lost its taxonomic significance, and to adopt as independent genera *Contortospiculum* Skrjabin, 1915 (with the species *Filaria labiata* Creplin, 1825), and *Molinema* Freitas et Lent, 1939 (with the species *Filaria bifida* Molin, 1858).

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Received
20.XII.1944.

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